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GENERAL SCIENCE
FOR TROPICAL SCHOOLS

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GENERAL SCIENCE

FOR TROPICAL SCHOOLS

BY

F. DANIEL

BOOK IV

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GENERAL SCIENCE FOR TROPICAL SCHOOLS

By F. DANIEL

A four-year course for secondary schools, especially written for use in tropical and sub-tropical areas. It consists of four books (I–IV), with an optional volume on Health Science (IIIA). There is also a Teacher's Handbook, which gives practical suggestions for the design and equipment of laboratories, lists of apparatus necessary for this General Science course, and useful hints arising out of the author's experience in conducting it.

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By F. DANIEL

A self-contained textbook *designed primarily for students not following the author's General Science course* and covering the syllabus in Health Science for the Cambridge Oversea School Certificate.

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BOOK II: LIMESTONE, SALT, AND THEIR DERIVATIVES: carbon dioxide—common salt—hydrogen chloride—chlorine—acids, bases, and salts.

PROPERTIES OF MATTER: surface tension—air pressure—liquid pressure.

PLANT LIFE: leaves—photo-synthesis—respiration—transpiration—importance of water. Flowers; pollination and fertilization. Fruits and seeds.

THE SOIL: formation—constituents—mineral matter—organic matter—air and water in the soil—plant food—micro-organisms—manures—composting—crop rotation.

SOUND: vibration—waves—resonance—musical instruments.

BOOK III: HEAT: temperature and thermometers—expansion—movement of heat—ice, water, and steam—humidity—quantity of heat—change of state—latent heat—heat and work—heat engines—heat and living things.

ANIMAL LIFE: the highest animals—food and digestion—transport—respiration—excretion—cells and tissues.

OTHER LIVING THINGS: the simplest living things—evolution—non-flowering plants—invertebrates—lower vertebrates.

LIGHT: nature and laws—reflection—refraction—lenses—optical instruments—colour—light and living things.

MAGNETISM: properties of magnets—the Earth's magnetism—electro-magnets.

APPENDIX: Practical work on animals.

BOOK IV: CHEMISTRY: chemical theory—common elements and their compounds—carbon and some of its compounds—food-stuffs.

PHYSIOLOGY: structure and functions of living matter; nutrition, respiration, excretion, growth, movement, in plants and animals; how living things behave; reproduction in plants and animals.

MECHANICS: force, work, and energy—machines—forces in equilibrium—motion—the measurement of time—the Earth in space.

ELECTRICITY: electric charges and electric currents—effects produced by electric currents—measurement of electricity—motors, dynamos, and power supplies—transformers—telegraphs and telephones—the nature of electricity.

APPENDIX: table of elements.

BOOK IIIA: (HEALTH SCIENCE): Healthy living; food and nutrition. HYGIENE: of the digestive system; of the respiratory system; of the blood system; of the skeletal, muscular, and nervous systems; of the skin. The conquest of disease. Public health. Food tables.

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The aims, scope, and teaching of General Science; accommodation, equipment, and laboratory management; list of apparatus, tools, and materials required; sketch-plans of science rooms; teaching methods; running notes on Books I–IV with additional experiments; bibliography; examination papers.

CONTENTS

BOOK FOUR

PREFACE TO THE SECOND EDITION xiv

SECTION I. CHEMISTRY

CHAPTER I. CHEMICAL THEORY 1

The atomic theory—the laws of chemical combination by weight—the law of conservation of matter—the law of fixed proportions—the law of multiple proportions—the law of combining weights—the chemical equivalents of elements (combining or equivalent weights)—valency—molecules—chemical symbols—chemical formulae—atomic weights—molecular weights—chemical equations—chemical equations for reactions studied already—Gay-Lussac's law of combining volumes—Avogadro's law—the standard volume (or the gram-molecular volume)—vapour density—oxidation and reduction—to find the composition of water by weight.

CHAPTER II. SOME IMPORTANT ELEMENTS AND THEIR COMPOUNDS 21

COMPOUNDS OF NITROGEN: ammonia—properties of ammonia—the nitrogen cycle—artificial fertilizers (chemical manures)—'fixing' atmospheric nitrogen—nitric acid—oxides of nitrogen—explosives—the extraction of Sicilian sulphur—the extraction of American sulphur—allotropic forms of sulphur—uses of sulphur. COMPOUNDS OF SULPHUR: hydrogen sulphide (or sulphuretted hydrogen) H_2S—properties of hydrogen sulphide—sulphur dioxide SO_2—properties of sulphur dioxide. SULPHURIC ACID: sulphur trioxide SO_3—the manufacture of sulphuric acid—the lead chamber process—the contact process. ALKALI MANUFACTURE: Solvay's process (the ammonia-soda process)—caustic alkalis (caustic soda and caustic potash). METALS AND METALLURGY: metals in everyday life—corrosion—alloys—mineral ores—general principles of extracting metals from ores. IRON AND STEEL: the metallurgy of iron—cast iron—wrought-iron—steel—steel manufacture—the properties of steel. SILICON COMPOUNDS IN EVERYDAY LIFE: glass—the properties of glass—clay, bricks, and pottery—cement—carborundum. PHOSPHORUS: the phosphorus cycle in Nature—yellow phosphorus—red phosphorus—matches—phosphorus pentoxide (P_2O_5) and phosphoric acid (H_3PO_4).

CHAPTER III. CARBON AND SOME OF ITS COMPOUNDS 55

Charcoal—lamp-black—animal charcoal. CRYSTALLINE FORMS OF CARBON. COAL. COAL-GAS: by-products from coal-gas manufacture. COMPOUNDS OF CARBON: carbon monoxide—properties of carbon monoxide—producer-gas and water-gas. HYDROCARBONS: methane—the paraffin series of hydrocarbons—the ethylene series of hydrocarbons—the acetylene series of hydrocarbons—the benzene series of hydrocarbons. PETROLEUM: petroleum refining. COMBUSTION AND FLAME: the candle flame—the Bunsen flame—the safety lamp. CARBOHYDRATES: complex sugars (di-saccharides)—simple sugars (mono-saccharides)—experiments with simple sugars (mono-saccharides)—experiments with complex sugars (di-saccharides)—complex carbohydrates (poly-saccharides)—experiments with complex carbohydrates (poly-saccharides)—to convert starch into simpler carbohydrates—glycogen—yeast—alcoholic fermentation—to show alcoholic fermentation of glucose—vinegar and acetic acid—citric acid and tartaric acid—'organic salts' (esters). OILS AND FATS: tests for fats—soaps—to prepare some hard soap—experiments with soap. PROTEINS: amino-acids—tests for proteins—Biuret reaction—nitric acid reaction—virus proteins.

SECTION II. PHYSIOLOGY

CHAPTER IV. THE STRUCTURE AND FUNCTIONS OF LIVING MATTER 97

Cells—cell structure—examination of cells—the functions of living protoplasm—nutrition and metabolism in living things.

CHAPTER V. NUTRITION IN PLANTS AND ANIMALS 102

NUTRITION IN PLANTS: nutrition in green plants—leaf structure—metabolism in the leaf—storage organs in plants—cellulose and lignin—nutrition in non-green plants. NUTRITION IN MAMMALS: nutrition in mammals (e.g. Man)—experiments illustrating digestion in the mouth—action of saliva—digestion in the stomach—experiment illustrating digestion in the stomach; action of pepsin—the clotting of milk by gastric juice—experiments with bile—experiments with emulsions—digestion in the small intestine—pancreatic juice—experiments illustrating pancreatic digestion—digestion in the small intestine; action of intestinal juice—the final products of digestion—absorption—diffusion, osmosis, and dialysis—experiments on osmosis and dialysis—assimilation. NUTRITION IN OTHER ANIMALS: nutrition in birds (e.g. pigeon)—nutrition in amphibians (e.g. frog)—nutrition in fishes—nutrition in molluscs (e.g. snail)—nutrition in insects (e.g. cockroach)—nutrition in worms (e.g. earthworm)—nutrition in stinging animals (e.g. Hydra)—nutrition in one-celled animals (e.g. Paramecium and Amoeba)—summary of nutrition in animals.

CHAPTER VI. RESPIRATION IN PLANTS AND ANIMALS 133

TISSUE RESPIRATION. RESPIRATORY MECHANISMS IN ANIMALS: respiration in Amoeba and Paramecium—respiration in Hydra—respiration in the earthworm—respiration in insects—respiration in the snail—respiration in fishes—respiration in the frog—respiration in birds (e.g. pigeon)—respiration in mammals (e.g. Man)—body-temperature (animal heat)—regulation of body-temperature—sweat—preventing loss of heat—food and energy—energy output and food requirements—energy values of foods. RESPIRATION IN PLANTS: respiration in green plants—in land plants—in submerged water plants—respiration in non-green plants—experiments on respiration. BLOOD: the clotting of blood—examination of heart and lungs.

CHAPTER VII. EXCRETION IN PLANTS AND ANIMALS 168

Excretion in animals—examination of kidney—excretion in green plants.

CHAPTER VIII. MOVEMENT AND BEHAVIOUR IN LIVING THINGS 173

Movement in plants—movements in response to gravity—plant movements in response to light—plant movements in response to water—movement in animals—skeletons and movement—teeth—kinds of movement in animals—muscles and their action—sensitivity, irritability and co-ordination—'plant behaviour'—animal behaviour—the nervous system of vertebrates—the sense-organs—sight and the eye—dissection of eye—hearing and the ear—experiments on hearing—sense of balance and position—internal sense-organs—senses of taste and smell—feeling, touch, temperature, and pain. Chemical control: hormones—instinct and intelligence.

CHAPTER IX. GROWTH AND REPRODUCTION IN LIVING THINGS 215

Growth of plants—in dry weight, in length, in thickness—growth-measurements. Reproduction in plants—vegetative reproduction in flowering plants. Sexual reproduction. Reproduction in lower animals—reproduction in mammals. CELL DIVISION: cell-division of sex-cells—inheritance of characteristics.

SECTION III. MECHANICS

CHAPTER X. MOTION AND FORCE 260

Motion and rest—speed and velocity—the composition of velocities—acceleration—the equations of motion—falling bodies—the acceleration due to gravity—Newton's First Law of Motion—inertia—force—forces

that produce no motion—the force of gravity—friction—the measurement of force—the addition of forces—inclined forces acting at a point—the triangle, parallelogram, and polygon of forces—the resolution of forces—parallel forces—equilibrium and stability.

CHAPTER XI. ENERGY 287

Work—power—potential energy—kinetic energy—water power—transformations of energy—the conservation of energy.

CHAPTER XII. MACHINES 298

Levers—pulleys and lifting-tackle—the principle of work—velocity ratio—mechanical advantage—efficiency—combinations of pulleys—the simple wheel and axle—the differential wheel and axle—the differential pulley-block—the inclined plane—the screw.

CHAPTER XIII. TIME AND SPACE 317

THE MEASUREMENT OF TIME—water-clocks—hour-candles—sand-glasses—sundials—the simple pendulum—experiments with pendulums—clocks and watches. THE EARTH IN SPACE—the heavens—the stars—the solar system—the Moon—planetoids, comets, and meteors.

SECTION IV. ELECTRICITY

CHAPTER XIV. ELECTRIC CHARGES AND ELECTRIC CURRENTS 333

STATIC ELECTRICITY—electricity produced by friction—positive and negative charges—conductors and non-conductors—simple electroscopes—electrostatic induction—lightning. CURRENT ELECTRICITY—the simple voltaic cell—the Leclanché cell—the dry cell—the lead accumulator—the screw rule.

CHAPTER XV. THE EFFECTS PRODUCED BY ELECTRIC CURRENTS 347

THE HEATING EFFECT OF AN ELECTRIC CURRENT—electric heaters—fuses—electric lighting—the arc lamp. THE CHEMICAL EFFECTS OF ELECTRIC CURRENTS—ions—electrolysis of copper sulphate—electro-plating—electrotypes—electrometallurgy—quantity of electricity. THE MAGNETIC EFFECTS OF ELECTRIC CURRENTS—the magnetic field round a straight conductor—the magnetic field round a circular conductor.

CHAPTER XVI. THE MEASUREMENT OF ELECTRICITY 363

CURRENT-MEASURING INSTRUMENTS: moving-iron instruments—moving-magnet instruments—moving-coil instruments—hot-wire instruments. ELECTRICAL UNITS: electromotive force—electric power—electrical resistance—Ohm's Law—ammeters and voltmeters—conductors in series and in parallel.

CHAPTER XVII. MOTORS, DYNAMOS, AND ELECTRIC POWER SUPPLIES 374

Electric motors—electro-magnetic induction—the magneto-electric effect—dynamos—the induction coil—the A.C. transformer—transmission of electrical power—conversion of A.C. to D.C.

CHAPTER XVIII. TELEGRAPHS AND TELEPHONES 388

CHAPTER XIX. THE NATURE OF ELECTRICITY AND THE STRUCTURE OF MATTER 392

TABLE OF ELEMENTS 398

GLOSSARY 399

INDEX 403

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PREFACE TO THE SECOND EDITION

During the fifteen years that have elapsed since these books were first published, General Science teaching has expanded rapidly, and half a million copies of these books have been used in oversea schools. During the same period there has been a steady advance in the qualifications and experience of oversea science teachers, in school science equipment, and in the educational attainments (particularly in English) of oversea students following this General Science Course. At the same time, science itself has progressed; new scientific facts have emerged, and some of the older theories have been modified. Minor amendments were made in successive reprintings of the first edition, but more extensive revision had to await the preparation of this new edition. The original textbooks have had a far-reaching influence on science teaching in oversea schools and it is hoped that this new edition will provide increasing numbers of oversea students with a better preparation for life in this scientific age.

F. DANIEL

Note.—The *Teacher's Handbook* that accompanies this General Science Course is designed to help the teacher to use the textbooks to the best advantage.

In the text, the asterisk sign (*) is used to mark the first appearance of a general word that is not included in the standard vocabulary on which the course is based (and which was not used in Books I, II, or III). Such words are defined in the glossary on p. 399.

The dagger sign (†) marks the first appearance of a new scientific word or technical term that has not been used in Books I, II, and III. Such words are also marked (†) in the Index so that the student can refer back to find where they were first used and explained in the text.

SECTION 1

CHEMISTRY

CHAPTER I

CHEMICAL THEORY

THE ATOMIC THEORY

When we studied the general characteristics of solids, liquids, and gases (Book Two, Chap. IV), we found that we could explain their physical behaviour by considering matter to be made up of tiny particles called *molecules*. Over 2,000 years ago some of the Greeks believed that if any lump of matter were divided and subdivided, again and again, into smaller and smaller pieces, the process of division would finally reach a limit, because matter was made up of exceedingly small, indivisible particles, which they called *atoms*. This *theory of atoms* was merely an intelligent guess, because the Greeks carried out no experiments to test the correctness of their ideas, and it is only in recent years that scientists have proved the actual existence of atoms and molecules and have measured their size and weight.

This early theory of atoms was of little practical use until about 1808, when *John Dalton* stated his *atomic theory* in such a way that it explained the experimental facts of chemical combination that were being discovered at this time. The chief points in *Dalton's Atomic Theory* that concern us in this Course are:

(1) *All matter is composed of exceedingly small particles called atoms.*

(2) *Atoms are indivisible during chemical reactions and can be neither created nor destroyed by chemical processes.*

(3) *All the atoms of the same element are exactly alike in every*

way (e.g. they all have the same individual weight) *but they are different from the atoms of other elements.*

(4) *Chemical compounds are formed by the combination of two or more atoms of different elements.*

(5) *Chemical combination takes place between simple whole numbers of atoms*, e.g. in the proportions 1:1, 1:2, 2:3, 3:4, etc.

Dalton's Atomic Theory is still accepted as the basis of chemical theory, although recent discoveries have shown that some of his statements now need slight modification. The theory is accepted because it explains, among other things, the *Laws of Chemical Combination by Weight.*

THE LAWS OF CHEMICAL COMBINATION BY WEIGHT

The Laws of Chemical Combination by Weight are:

(i) *The Law of Conservation of Matter* (or the Law of Indestructibility* of Matter).

(ii) *The Law of Constant Composition* (or the Law of Fixed Proportions).

(iii) *The Law of Multiple Proportions.*

(iv) *The Law of Combining Weights* (also called the Law of Equivalents or the Law of Reciprocal* Proportions).

All these chemical laws, which are based on the results of accurate experiments, can be explained by Dalton's Atomic Theory, and this is why his theory is so important.

THE LAW OF CONSERVATION OF MATTER

The French chemist, Lavoisier, from experiments carried out in 1774, concluded that *matter can be neither created nor destroyed during chemical reactions.* Just as the Law of Conservation of Energy is the basis of all quantitative* physics, this Law of Conservation of Matter is the basis of all quantitative chemistry, since it follows that *in all chemical changes the total weight of all the reacting substances is exactly the same as the total weight of all the products of the reaction.*

This law is explained by Dalton's suggestion that atoms can be

neither created nor destroyed *in chemical reactions*, and that chemical changes are merely regroupings of atoms to form different molecules.

THE LAW OF CONSTANT COMPOSITION

About 1800 the French chemist, Proust, after carrying out a series of quantitative experiments, stated that *the same chemical compound always consists of the same elements combined together in the same proportions by weight*. This is the *Law of Constant Composition* or the *Law of Fixed Proportions*.

This fact that a pure chemical compound always has the same composition by weight is explained by Dalton's suggestion that all the atoms of the same element are exactly alike in weight. Since every molecule of a chemical compound always consists of the same numbers and kinds of atoms, if the weight of each atom is fixed then the compound must have a constant composition. For example, it is always found that pure water contains 8 parts by weight of oxygen combined with 1 part by weight of hydrogen, no matter what the source of the water may be. It is the same with every other pure substance known to scientists, e.g. whether we make common salt by burning sodium in chlorine, by neutralizing caustic soda with hydrochloric acid, or by dissolving sodium carbonate in hydrochloric acid: the sodium chloride produced will always contain sodium and chlorine in exactly the same proportions by weight. This is because the smallest possible particle of common salt—*one molecule*—contains *one atom* of sodium combined with *one atom* of chlorine; and since all the sodium atoms are of the same weight, and similarly all the chlorine atoms, the proportion of sodium to chlorine must be always the same. (We consider *air* to be a *mixture* and not a *compound* of nitrogen and oxygen, because the relative proportions of nitrogen and oxygen in the air *are not always exactly the same* at different times and places.)

THE LAW OF MULTIPLE PROPORTIONS

Dalton concluded from his Atomic Theory that *when two ele-*

ments (A and B) combine to form more than one compound, the different weights of one (B) that combine with a fixed weight of the other (A) bear a simple ratio to one another. This is the *Law of Multiple Proportions.*

The correctness of this law was afterwards shown by careful analysis of a large number of compounds. The law agrees with Dalton's suggestion that combination takes place between simple whole numbers of atoms, and since one atom of an element A combines with one or two or three or more atoms of an element B, the different weights of B that combine with a fixed weight of A will bear a simple ratio to each other, e.g. 1:2; 1:3; 2:3; 3:4; etc.

The same two elements frequently combine to form more than one compound, e.g. there are two oxides of carbon: (*a*) *carbon monoxide*,† formed when carbon burns in a limited supply of air, and (*b*) *carbon dioxide*, formed when carbon burns in a free supply of air. Analysis shows that these two oxides of carbon have the following composition:

	Carbon monoxide	Carbon dioxide
Carbon	1·000	1·000
Oxygen	1·333	2·666

Hence, in *carbon dioxide* there is exactly *twice* as much oxygen combined with one part of carbon as there is in *carbon monoxide*: so in the two oxides of carbon the weights of oxygen that combine with the same weight of carbon bear a simple ratio to one another, i.e. 1:2. This is because a molecule of carbon monoxide contains one atom of carbon combined with one atom of oxygen, while a molecule of carbon dioxide contains one atom of carbon combined with two atoms of oxygen.

To take another example, there are two oxides of sulphur, (*a*) *sulphur dioxide*, formed when sulphur burns in air, and (*b*) *sulphur trioxide*,† formed when sulphur dioxide is oxidized (i.e. when it combines with more oxygen). Analysis shows that these two compounds have the following composition:

	Sulphur dioxide	Sulphur trioxide
Sulphur	2·000	2·000
Oxygen	2·000	3·000

That is, in the two oxides of sulphur, the weights of oxygen that combine with the same weight of sulphur bear a simple ratio to one another, i.e. 2:3. This is because a molecule of sulphur dioxide contains one atom of sulphur combined with two atoms of oxygen, while a molecule of sulphur trioxide contains one atom of sulphur combined with three atoms of oxygen.

Similarly, there are two oxides of copper: (*a*) *cuprous oxide*† (formed as a red precipitate when testing for sugar with Fehling's solution), and *cupric oxide*† (the ordinary black oxide). These two oxides have the following composition:

	Cuprous oxide	Cupric oxide
Copper	1·000	1·000
Oxygen	0·126	0·252

That is, in *cupric* oxide there is *twice* as much oxygen as there is in *cuprous* oxide, or the weights of oxygen that combine with the same weight of copper bear a simple ratio to one another, i.e. 2:1. This is because a molecule of cupric oxide contains one atom of copper combined with one atom of oxygen, while a molecule of cuprous oxide contains two atoms of copper combined with one atom of oxygen. (The ending *-ous* is added to the Latin name of an element to show that the compound contains a smaller proportion of another element. The ending *-ic* shows the higher proportion of another element. For example, cupric oxide contains a higher proportion of oxygen than cuprous oxide. Similarly, sulphuric acid contains a higher proportion of oxygen than sulphurous acid.†)

THE LAW OF COMBINING WEIGHTS

The weights of the elements A and B that combine with a fixed weight of another element C are either the weights in which A and B combine together or simple multiples of these weights. This law follows from the Law of Constant Proportions and the Law of Multiple Proportions. It is explained by Dalton's suggestions (3), (4), and (5) on pp. 1–2.

Some actual examples will make this law clear:

23 gm. of *sodium* combine with 1 gm. of *hydrogen* to form *sodium hydride.*†

35·5 gm. of *chlorine* combine with 1 gm. of *hydrogen* to form *hydrogen chloride.*

23 gm. of *sodium* combine with 35·5 gm. of *chlorine* to form *sodium chloride.*

Fig. 1 illustrates this relationship between the combining weights of hydrogen, sodium, and chlorine.

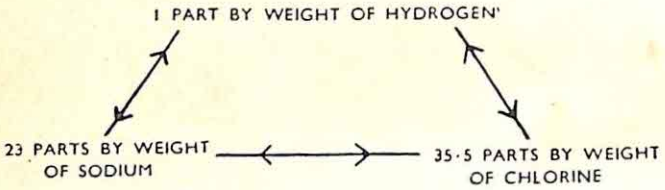


FIG. 1. Equivalent proportions.

Hence the weights of sodium and chlorine that combine separately with one part by weight of hydrogen are also the weights in which sodium and chlorine combine with each other.

Similarly,

1 gm. of *hydrogen* combines with 8 gm. of *oxygen* to form *water.*

1 gm. of *hydrogen* combines with 3 gm. of *carbon* to form *methane.*†

3 gm. of *carbon* combine with 8 gm. of *oxygen* to form *carbon dioxide.*

Hence the weights of carbon and oxygen that combine separately with one part by weight of hydrogen are also the weights in which carbon and oxygen combine with each other.

THE CHEMICAL EQUIVALENTS OF ELEMENTS (COMBINING OR EQUIVALENT WEIGHTS)

The Law of Combining Weights leads to the idea that each element has a definite *chemical equivalent.* In the above examples it is clear that 1 *part of hydrogen,* 23 *parts of sodium,* 35·5 *parts of*

chlorine, 8 parts of oxygen, and 3 parts of carbon are equal to each other in combining power, i.e. they are *chemical equivalents*. It is convenient to choose a standard weight of a suitable substance in terms of which the *chemical equivalents* (or *equivalent weights*) of all elements (and compounds) can be expressed, and *one unit weight of hydrogen* is taken as this standard.

Hence, when we say that the *equivalent weight of sodium* is 23, we mean that 23 gm. of sodium will combine with 1 gm. of hydrogen, or will displace 1 gm. of hydrogen from its compounds.

The chemical equivalent (or equivalent weight) of an element is that weight of it which combines with, or takes the place of, one part by weight of hydrogen.

Notice that the chemical equivalent of an element is not a weight but merely a *ratio*, i.e a number.

The equivalent weight of an element can be found experimentally with great accuracy, and this forms the starting-point for all quantitative work in chemistry. Your teacher will show you how to find the equivalent weight of a few common elements in the laboratory.

This principle of *chemical equivalents* (or *combining weights*) applies also to chemical compounds, e.g. acids and bases. The *equivalent weight* (or *gram-equivalent*†) of an acid is the number of grams of the acid that contain one gram of hydrogen replaceable by a metal. The *equivalent weight* (or *gram-equivalent*) of a base is the number of grams of the base required to neutralize one gram-equivalent of an acid.

A knowledge of equivalent weights enables us to prepare the *standard solutions* used for *volumetric analysis*,† in which we measure the *volumes* of the reacting substances (instead of their weights—as in *gravimetric analysis*†). A *normal solution*† of a substance is one that contains *one gram-equivalent of the substance per litre of solution*. Hence a litre of normal hydrochloric acid solution exactly neutralizes a litre of normal sodium hydroxide solution. So if we want to find the amount of sodium hydroxide present in a litre of our bench reagent, we measure out a known

volume of this solution (using a *pipette*), add an *indicator* (e.g. litmus), and then run into it (from a *burette*) our standard solution of hydrochloric acid until the litmus becomes purple, showing that the liquid is *neutral* and that the reaction is, therefore, complete. (See *General Science Workbook*, Book Two, pp. 155–7, for numerical problems on volumetric analysis.)

VALENCY

The valency† of an element is the number of hydrogen (or chlorine) atoms that it will combine with or displace.

For example, one atom of sodium can only combine with or displace *one* atom of hydrogen or chlorine; hence sodium is said to have a valency of one. One atom of zinc, however, can combine with *two* chlorine atoms; hence zinc has a valency of two. One atom of aluminium can combine with three chlorine atoms; hence the valency of aluminium is three. In the same way, carbon has a valency of four, phosphorus has a valency of five, and sulphur has a valency of six. No element has a higher valency than eight.

Some elements show more than one valency and so have more than one equivalent weight, e.g. sulphur shows a valency of 4 (and an equivalent weight of 8) in sulphur dioxide, and a valency of 6 (and an equivalent weight of 5·3) in sulphur trioxide. Copper has a valency of 1 (and an equivalent weight of 63·6) in cuprous oxide, and a valency of 2 (and an equivalent weight of 31·8) in cupric oxide.

We shall find that there is a simple relation between the *valency*, the *equivalent weight*, and the *atomic weight* of an element, thus:

$$\text{VALENCY} = \frac{\text{ATOMIC WEIGHT}}{\text{EQUIVALENT WEIGHT}}.$$

This gives us another definition of *valency*: *The number of chemical equivalents in the atomic weight of the element.*

A knowledge of valencies will be very useful when we come to use chemical formulae.

MOLECULES

An atom is the smallest particle of an element that can take part in

a chemical change. Very few atoms, however, can exist singly, by themselves, in a free state, but they usually combine together to form *molecules*.

A molecule is the smallest particle of an element or compound that can exist in a free state.

For example, one *atom* of sodium combines with one *atom* of chlorine to form one *molecule* of common salt. One atom of oxygen combines with two atoms of hydrogen to form one molecule of water. One atom of carbon combines with two atoms of oxygen to form one molecule of carbon dioxide. If the atoms in a molecule are of different kinds, then the substance is a *compound*. If the atoms in a molecule are all of the same kind, then the substance is an *element*, e.g. hydrogen and oxygen *molecules* each contain *two atoms*.

CHEMICAL SYMBOLS

Every element has a name, and, for convenience, a *symbol*,* which is usually the first letter of its English or Latin name, e.g. the symbol for *carbon* is C, *nitrogen* is N, *sulphur* is S, *hydrogen* is H, *iodine* is I, *oxygen* is O, *phosphorus* is P, *potassium* (Kalium) is K.

When the names of several elements all begin with the same letter, two letters are used, e.g. *chlorine*—Cl, *calcium*—Ca, *cobalt*—Co, *magnesium*—Mg, *manganese*—Mn, *nickel*—Ni, *zinc*—Zn, *platinum*—Pt, *aluminium*—Al, *sodium* (Natrium)—Na, *iron* (Ferrum)—Fe, *copper* (Cuprum)—Cu, *silver* (Argentum)—Ag, *gold* (Aurum)—Au, *antimony*† (Stibium)—Sb, *tin* (Stannum)—Sn, *mercury* (Hydrargyrum)—Hg.

It is to be remembered that these symbols should be used *only in chemical equations*. They are *not* shorthand signs for the elements, e.g. Pb does *not* stand for lead in general, any more than $1 means 'money' or 'wealth'. The symbol Pb stands for *one atom of lead*. On p. 11 you will find a list of the commoner chemical elements and their symbols.

CHEMICAL FORMULAE

The next step is to make use of these atomic symbols to represent

molecules, e.g. a molecule of oxygen contains two atoms of oxygen, and its *formula* is written O_2. Similarly, Cl_2 stands for a molecule of chlorine. (2Cl stands for two atoms of chlorine not combined together. $2Cl_2$ stands for two molecules of chlorine, each consisting of two atoms combined together.) NaCl stands for one molecule of sodium chloride containing an atom of sodium combined with an atom of chlorine. $CaCO_3$ stands for one molecule of calcium carbonate, consisting of one atom of calcium combined with one atom of carbon and three atoms of oxygen. A molecule of *carbon monoxide* contains one atom of carbon combined with one atom of oxygen; therefore its formula is CO. A molecule of *carbon dioxide* contains one atom of carbon combined with two atoms of oxygen; therefore its formula is CO_2. Your teacher will show you how to make up chemical formulae for the chemical compounds you have met in this Course.

ATOMIC WEIGHTS

According to Dalton's Atomic Theory, all the atoms of an element are equal in weight, but differ in weight from the atoms of other elements. A single atom, however, is so exceedingly small that the earlier chemists made no attempt to weigh *one atom*; but they found the *relative atomic weights* of the elements, finding *how many times an atom of each element was heavier than an atom of hydrogen* (the lightest element). The atomic weight of an element is therefore (like its chemical equivalent), merely a *ratio*:

$$\text{ATOMIC WEIGHT} = \frac{\text{WEIGHT OF 1 ATOM OF ELEMENT}}{\text{WEIGHT OF 1 ATOM OF HYDROGEN}}.$$

For example, the oxygen atom is 16 times heavier than the hydrogen atom, so that if the *atomic weight of hydrogen* = 1, then the *atomic weight of oxygen* = 16. When we say, for example, that the atomic weight of sulphur is 32, we do not mean that we have actually weighed an atom of sulphur, but that an atom of sulphur is 32 times heavier than an atom of hydrogen, or twice as heavy as an atom of oxygen. A list of the atomic weights of some of the

commoner elements is given below. A complete list will be found on p. 398.

We have learnt that there is a simple relation between the *valency*, the *equivalent weight*, and the *atomic weight* of an element, and that the atomic weight is either the same as the equivalent

English Name	Latin Name	Symbol	Equivalent Weight	Valency	Atomic Weight
ALUMINIUM		Al	9	3	27
CALCIUM		Ca	20	2	40
CARBON		C	3	4	12
COPPER	CUPRUM	Cu	31·8 or 63·6	1 or 2	63·6
CHLORINE		Cl	35·5	1	35·5
HYDROGEN		H	1	1	1
IRON	FERRUM	Fe	28 or 18·67	2 or 3	56
LEAD	PLUMBUM	Pb	207, 103·5, or 51·75	1, 2, or 4	207
MAGNESIUM		Mg	12·15	2	24·3
MERCURY	HYDRAR-GYRUM	Hg	200·6 or 100·3	1 or 2	200·6
NITROGEN		N	4·67 or 2·8	3 or 5	14
OXYGEN		O	8	2	16
PHOSPHORUS		P	10·33 or 6·2	3 or 5	31
POTASSIUM	KALIUM	K	39	1	39
SILICON		Si	7	4	28
SODIUM	NATRIUM	Na	23	1	23
SULPHUR		S	16, 8, or 5·33	2, 4, or 6	32
TIN	STANNUM	Sn	59·5 or 29·75	2 or 4	119
ZINC		Zn	32·7	2	65·4

weight or is some simple multiple of it, i.e. *atomic weight* = *equivalent weight* × *valency*. Atomic weights, therefore, are fixed by first finding the equivalent weight very accurately, and then finding the valency (which is always a simple whole number between 1 and 8) and multiplying the two together.

MOLECULAR WEIGHTS

The *molecular weight* of a compound is simply the sum of the atomic weights of the atoms in one molecule of the compound; e.g. since the atomic weight of sodium is 23 and the atomic weight of chlorine is 35·5, the molecular weight of sodium chloride, NaCl, is $23+35{\cdot}5 = 58{\cdot}5$. To take another example, the formula of sulphuric acid is H_2SO_4: since H = 1, S = 32, and O = 16, the molecular weight of sulphuric acid is $(1\times2)+32+(16\times4) = 98$. This can be expressed in another way: *The molecular weight of a compound is the number of times that a molecule of the compound is heavier than one ATOM of hydrogen*, i.e.

$$\text{MOLECULAR WEIGHT} = \frac{\text{WEIGHT OF ONE MOLECULE OF THE COMPOUND}}{\text{WEIGHT OF ONE ATOM OF HYDROGEN}}.$$

CHEMICAL EQUATIONS

Just as we can use formulae to represent atoms and molecules, we can use *chemical equations* to represent *chemical changes* or *chemical reactions*. For example, when zinc dissolves in dilute sulphuric acid, hydrogen is given off, and zinc sulphate remains behind. We can represent this reaction as follows:

$$Zn + H_2SO_4 \longrightarrow ZnSO_4 + H_2$$

This equation tells us (*a*) *that under certain conditions, not discoverable from the equation, zinc reacts with sulphuric acid so as to form zinc sulphate and hydrogen*;

(*b*) *that one molecule of zinc reacts with one molecule of sulphuric acid to give one molecule of zinc sulphate and one molecule of hydrogen*;

(*c*) *knowing the atomic weights*, Zn = 65, H = 1, O = 16, S = 32, *the equation tells us that* 65 *gm., lb., or tons of zinc react with* 98 *gm., etc., of sulphuric acid, to give* 161 *gm., etc., of zinc sulphate and* 2 *gm., etc., of hydrogen*:

$$\underset{65}{Zn} + \underset{2+32+64\ =\ 98}{H_2SO_4} \longrightarrow \underset{65+32+64\ =\ 161}{ZnSO_4} + \underset{2}{H_2}$$

(Hence, if we know the atomic weights, we can work out the quantities of substances taking part in a chemical change.)

This equation does not tell us anything about the conditions under which the reaction takes place, e.g. whether it is necessary to heat or not, whether the acid is concentrated or dilute, what is the speed of the reaction, or whether a catalyst is necessary. Hence, in studying chemical reactions it is not enough merely to remember equations; they tell us a lot, but not everything.

In such equations, either an arrow (→) or the equation sign (=) may be used. The arrow is more correct because it indicates that a reaction is going on, usually in one particular direction, but the equation sign is sometimes preferred because it emphasizes that a balance exists between the number of atoms of each element on both sides of the equation.

CHEMICAL EQUATIONS FOR REACTIONS WE HAVE STUDIED ALREADY[1]

When iron combines with sulphur, on heating, the equation is:

$$Fe+S = FeS$$

The following equations represent the burning of various elements in air or oxygen:

$S+O_2 = SO_2$ $C+O_2 = CO_2$ $2Mg+O_2 = 2MgO$ $2Ca+O_2 = 2CaO$
$4P+5O_2 = 2P_2O_5$ $3Fe+2O_2 = Fe_3O_4$ (magnetic oxide of iron)

The equation that represents the preparation of oxygen from potassium chlorate by heating is $2KClO_3 = 2KCl+3O_2$. If manganese dioxide is used as a catalyst, it is not necessary to mention it in the equation because it remains unchanged at the end of the reaction, and would therefore appear on both sides of the equation and thus 'cancel out'.*

$2H_2O \longrightarrow 2H_2+O_2$ represents the *analysis* of water, while $2H_2+O_2 \longrightarrow 2H_2O$ represents the *synthesis* of water. If we want to show both these reactions in one equation, we write $2H_2+O_2 \rightleftarrows 2H_2O$, which means that the change is *reversible*.

[1] Copy these equations into Books One and Two.

The following equations are examples of the *displacement* of one element by another:

$Fe+CuSO_4 = FeSO_4+Cu$ (Displacement of copper from copper sulphate by iron)
$Zn+CuSO_4 = ZnSO_4+Cu$ (Displacement of copper from copper sulphate by zinc)
$Zn+H_2SO_4 = ZnSO_4+H_2$ (Displacement of hydrogen from sulphuric acid by zinc)
$Fe+H_2SO_4 = FeSO_4+H_2$ (Displacement of hydrogen from sulphuric acid by iron)
$Mg+H_2SO_4 = MgSO_4+H_2$ (Displacement of hydrogen from sulphuric acid by magnesium)
$Zn+2HCl = ZnCl_2+H_2$ (Displacement of hydrogen from hydrochloric acid by zinc)
$Fe+2HCl = FeCl_2+H_2$ (Displacement of hydrogen from hydrochloric acid by iron)
$2Na+2H_2O = 2NaOH+H_2$ (The action of sodium on water)
$2K+2H_2O = 2KOH+H_2$ (The action of potassium on water)
$Ca+2H_2O = Ca(OH)_2+H_2$ (The action of calcium on water)
$Mg+H_2O = MgO+H_2$ (The action of magnesium on steam)
$3Fe+4H_2O = Fe_3O_4+4H_2$ (The action of steam on red-hot iron)

The following further examples of chemical equations are for reactions studied in Book Two:

$CaCO_3 = CaO+CO_2$
$CaCO_3+2HCl = CaCl_2+H_2O+CO_2$
$4HCl+MnO_2 = MnCl_2+2H_2O+Cl_2$
$KOH+HNO_3 = KNO_3+H_2O$
$ZnCO_3+H_2SO_4 = ZnSO_4+CO_2+H_2O$
$CaO+H_2O = Ca(OH)_2$
$CaCO_3+H_2O+CO_2 \rightleftarrows Ca(HCO_3)_2$
$NaOH+HCl = NaCl+H_2O$
$PbO+2HNO_3 = Pb(NO_3)_2+H_2O$
$CuO+H_2SO_4 = CuSO_4+H_2O$

GAY-LUSSAC'S LAW OF COMBINING VOLUMES

When we study chemical reactions between *gases* we find that *gases always combine together in simple proportions by volume.* For example, 2 volumes of hydrogen combine with 1 volume of oxygen to form 2 volumes of steam; 1 volume of hydrogen combines with 1 volume of chlorine to form 2 volumes of hydrogen chloride; 1 volume of nitrogen combines with 3 volumes of hydrogen to form 2 volumes of ammonia gas. All gases behave in a similar way when they combine, one volume of a gas combining with the same, or twice, or three or four times this volume of another gas, but never in odd amounts. The French chemist, Gay-Lussac, studied a large number of reactions between gases, and, about 1808, he came to the conclusion that *when gases combine, they do so in a simple ratio by volume to each other, and to the gaseous product.* This is known as *Gay-Lussac's Law of Combining Volumes.*

AVOGADRO'S LAW

The simple relation that exists between the volumes of combining

gases is explained by Avogadro's Law, which states that *equal volumes of all gases* (under the same conditions of temperature and pressure) *contain the same number of molecules.* This theory was first suggested by the Italian chemist, Avogadro, about 1811, and it gives us a clearer idea of how chemical reactions take place. For example, in the combination of hydrogen with chlorine, 1 volume of hydrogen combines with 1 volume of chlorine to form 2 volumes of hydrogen chloride. But, according to Avogadro's Law, equal volumes of gases contain *the same number of molecules*; therefore *one molecule* of hydrogen combines with *one molecule* of chlorine to form *two molecules* of hydrogen chloride (see Fig. 2).

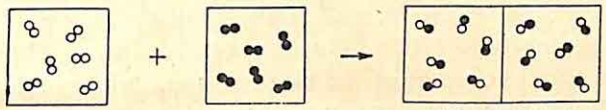


FIG. 2. Combining volumes of gases.

Hence, one molecule of hydrogen chloride must contain *half a molecule* of hydrogen and *half a molecule* of chlorine, i.e. the *molecules* of hydrogen and chlorine must each contain at least *two atoms.* Avogadro's Law connects together Dalton's Atomic Theory and Gay-Lussac's Law of Volumes.

We see, therefore, that our chemical equations also tell us the *volumes* of the gases taking part in chemical changes, for example:

$$\underset{1\text{ vol.}}{N_2} + \underset{3\text{ vols.}}{3H_2} \longrightarrow \underset{2\text{ vols.}}{2NH_3} \quad \Big| \quad \underset{1\text{ vol.}}{H_2} + \underset{1\text{ vol.}}{Cl_2} \longrightarrow \underset{2\text{ vols.}}{2HCl} \quad \Big| \quad \underset{2\text{ vols.}}{2H_2} + \underset{1\text{ vol.}}{O_2} \longrightarrow \underset{2\text{ vols.}}{2H_2O}$$

THE STANDARD VOLUME (OR THE GRAM-MOLECULAR VOLUME)

Since equal volumes of all gases contain equal numbers of molecules, the weights of these equal volumes of gases will be proportional to the *molecular weights* of the gases. Hence, *the molecular weight of any gas, in grams, occupies the same volume* (22·2 litres) *at normal temperature and pressure* (0° C. and 760 mm. of mercury).

(By experiment, a litre of hydrogen weighs 0·09 gm. at N.T.P.

But the molecular weight of hydrogen is 2. So 2 gm. of hydrogen occupy $\frac{1\times 2}{0{\cdot}09} = 22{\cdot}2$ litres, at normal temperature and pressure. Hence, by Avogadro's Law, *the gram-molecular volume of any gas is* 22·2 *litres at N.T.P.*)

This fact is very useful when doing chemical calculations. For example, in the reaction $Zn+H_2SO_4 \longrightarrow ZnSO_4+H_2$, we can now say that 65 gm. of zinc react with 98 gm. of sulphuric acid to give 161 gm. of zinc sulphate and 2 gm. *or* 22·2 *litres of hydrogen* (*at N.T.P.*). *Hence a chemical equation also tells us the volumes of gases* (*at N.T.P.*) *taking part in the reaction.*

The *vapour density*† of a gas is the ratio of the weight of a certain volume of the gas to the weight of an equal volume of hydrogen (measured under the same conditions of temperature and pressure). Since these equal volumes of the two gases contain equal numbers of molecules, the vapour density of the first gas tells us how many times a molecule of this gas is heavier than a molecule of hydrogen. But the molecular weight of hydrogen is 2, so *the molecular weight of any gas is twice its vapour density.*

OXIDATION AND REDUCTION

When a substance combines with oxygen, it is said to be *oxidized.* The reverse change, when oxygen is removed from a substance, is called *reduction.*† For example, when lead is heated in air, it forms lead oxide:

$$2Pb+O_2 \longrightarrow 2PbO$$

This is the *oxidation* of lead to lead oxide. On the other hand, when lead oxide is heated on a charcoal block (with the blowpipe flame), oxygen is removed from the lead oxide, and metallic lead is left:

$$PbO+C \longrightarrow Pb+CO$$

This is the *reduction* of lead oxide to lead, and the carbon is called a *reducing agent.* Hydrogen is a common reducing agent, e.g.:

$$CuO+H_2 \longrightarrow Cu+H_2O$$

In some cases the hydrogen merely adds on to the other substance without actually removing oxygen, but the change is still *reduction*; e.g. when hydrogen and sulphur are heated together, some hydrogen sulphide (or sulphuretted hydrogen)† is formed:

$$H_2 + S = H_2S$$

Most oxidizing agents are substances that are rich in oxygen and that will part with it readily, e.g. potassium nitrate—KNO_3, nitric acid—HNO_3, manganese dioxide—MnO_2, potassium permanganate†—$KMnO_4$, potassium dichromate†—$K_2Cr_2O_7$, hydrogen peroxide—H_2O_2, lead peroxide—PbO_2.

Reducing agents are substances with a great chemical attraction for oxygen, e.g. hydrogen, carbon, carbon monoxide—CO, sulphur dioxide—SO_2.

Oxidation can take place in three ways:

(*a*) *by direct addition of oxygen to an element or compound*, e.g. the burning of sulphur in air to form sulphur dioxide;

(*b*) *by removal of hydrogen from a compound containing this element*, e.g. the incomplete combustion of hydrogen sulphide to form sulphur and water-vapour (see pp. 33–34);

(*c*) *by increasing the valency of an element* (even though there is no direct addition of oxygen or removal of hydrogen), e.g. in changing *ferrous* sulphate,† $FeSO_4$, to *ferric* sulphate,† $Fe_2(SO_4)_3$, when the valency of iron is *increased* from two to three.

Reduction (the converse change) can also take place in three ways:

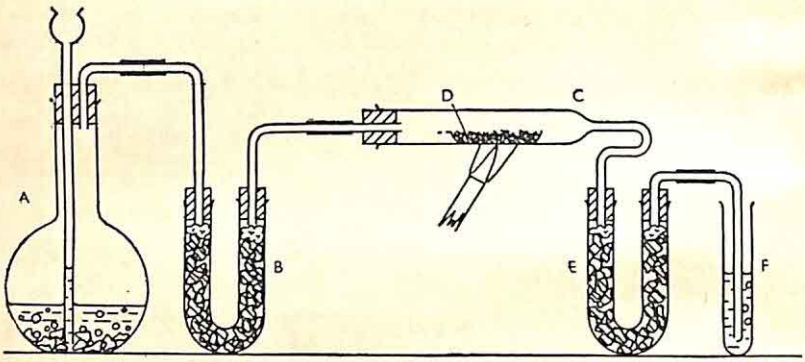
(*a*) *by removal of oxygen from an oxide*, e.g. when lead oxide is heated on a charcoal block;

(*b*) *by direct addition of hydrogen to an element or compound*, e.g. when hydrogen combines directly with sulphur to form hydrogen sulphide;

(*c*) *by decreasing the valency of an element*, e.g. in changing *ferric* sulphate, $Fe_2(SO_4)_3$, to *ferrous* sulphate, $FeSO_4$, when the valency of iron is *reduced* from three to two.

TO FIND THE COMPOSITION OF WATER BY WEIGHT

If we heat black copper oxide (cupric oxide) in a stream of hydrogen, the oxygen of the copper oxide combines with the hydrogen to form water, leaving metallic copper. In other words, the hydrogen *reduces* the copper oxide to copper, while the oxygen of the copper oxide *oxidizes* the hydrogen to water. This reaction gives us a convenient method for finding the composition of water by weight. It also enables us to find the *chemical equivalent* (or



(a)

FIG. 3. (*a*) Finding the composition of water by weight.

equivalent weight) of oxygen and the equivalent weight of copper in one single experiment:

$$CuO + H_2 \longrightarrow Cu + H_2O$$

Using the apparatus shown in Fig. 3, hydrogen is prepared from zinc and dilute sulphuric acid in the vessel A, and is then dried by passing through the drying-tube B, which contains fused calcium chloride. This dry hydrogen then passes through the hard-glass tube C and over some weighed copper oxide contained in the 'boat' D. When tube C is heated, the hydrogen reduces the hot copper oxide to metallic copper, and the water that is formed is caught in the second drying-tube E, which also contains fused calcium chloride. The increase in weight of tubes C and E during the experiment gives the amount of *water* formed. The loss in

weight of the boat D, containing copper oxide, gives the weight of *oxygen* contained in this weight of water. The difference between the weight of the water and that of the oxygen gives the weight of the *hydrogen* combined with oxygen in this water.

First dry the boat by heating it in the flame, holding it with crucible tongs; allow it to cool in the desiccator and weigh it (to the nearest 0·001 gm.), recording the weight at (*a*) on p. 20. Half-fill the boat with black copper oxide (prepared from thin copper wire), dry it by heating for a few minutes, allow it to cool in the desiccator, then weigh and record at (*b*). Also weigh the tubes C and E, with their corks and glass tubes, hanging them on the balance by means of a wire hook (see Fig. 3(*b*)). Record this weight at (*r*). From the drying-tube E, a glass tube dips into about $\frac{1}{2}$ in. of concentrated sulphuric acid in a test-tube F. This stops water-vapour in the atmosphere from getting into the drying-tube E, and the rate of bubbling also shows the speed at which hydrogen is passing through the apparatus. It also removes the risk of the excess hydrogen catching fire and causing an explosion. (If hydrogen is being given off in A and no bubbles appear in F, then your apparatus is not air-tight.)

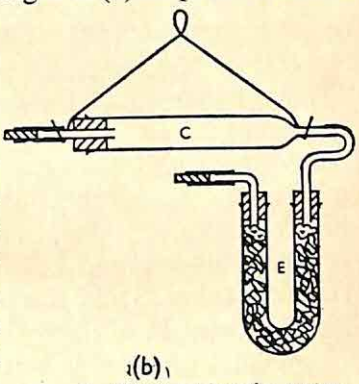


FIG. 3. (*b*) Weighing the water formed.

Pour about 100 c.cm. of dilute sulphuric acid (6N) on to the zinc in A (after adding a little copper sulphate solution to act as a catalyst) so as to give a steady stream of hydrogen over a long period. (To be on the safe side, cover the vessel A and the drying-tube B with a duster in case of an explosion.) When you think that all the air has been displaced from the apparatus, collect a small test-tube full of the gas coming through and light it. If it burns quietly, light your burner and heat the hard-glass tube C, gently at first, keeping the flame moving about all the time. When all

the black copper oxide appears to have been reduced to reddish-brown metallic copper, stop heating, and allow the copper to cool *in a stream of hydrogen.* When the hard-glass tube C is cool, remove the boat (with a wire hook), weigh it and record at (*c*). Then draw dry air through the tubes C and E and weigh them as before—(*s*).

Weight of empty 'boat' . . . . = ...gm. (*a*)
Weight of boat+copper oxide *before* experiment . . . . . = ...gm. (*b*)
Weight of boat+metallic copper *after* experiment . . . . . = ...gm. (*c*)
Therefore, weight of copper oxide (*b*—*a*) . = ...gm. (*d*)
Therefore, weight of metallic copper (*c*—*a*) = ...gm. (*e*)
Therefore, weight of oxygen (*d*—*e*) or (*b*—*c*) = ...gm. (*f*)
Weight of tubes C and E *before* experiment = ...gm. (*r*)
Weight of tubes C and E *after* experiment . = ...gm. (*s*)
Therefore, weight of *water* formed (*s*—*r*) . = ...gm. (*t*)
But, weight of *oxygen* in this water (*b*—*c*) . = ...gm. (*f*)
And, weight of *hydrogen* in this water (*t*—*f*) = ...gm.
Therefore, ...gm. of water contain ..... gm. of hydrogen and .....gm. of oxygen.
Therefore, 100 gm. of water contain .....gm. of hydrogen and .....gm. of oxygen.[1]

Your experiment also shows that .....gm. of oxygen combines with .....gm. copper. Therefore, 8 gm. of oxygen combine with .....gm. of copper. Since 8 gm. of oxygen combine with 1 gm. of hydrogen, the chemical equivalent of copper is .....

(See *General Science Workbook*, Book Two, pp. 148–57, for numerical problems and learning exercises on this chapter.)

[1] The most careful experiments show that water contains 11·1% of hydrogen and 88·9% of oxygen, i.e. 8 gm. of oxygen combine with 1 gm. of hydrogen, or the chemical equivalent of oxygen is 8.

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Chapter II

SOME IMPORTANT ELEMENTS AND THEIR COMPOUNDS

There are over ninety elements in the Earth's crust and at least one-third of them are important to Man. In this Course, however, we can study in detail only a few of these—*nitrogen*, *sulphur*, and *carbon*—but other elements and compounds are mentioned elsewhere in the Course.

COMPOUNDS OF NITROGEN

AMMONIA

Ammonia can be prepared by heating any *ammonium salt* with an alkali. Ammonium chloride and calcium hydroxide are generally used for this purpose in the laboratory:

$$2NH_4Cl + Ca(OH)_2 \longrightarrow CaCl_2 + 2H_2O + 2NH_3$$

Ammonia is a gas that is very soluble in water; hence it cannot be collected over water. As it is much 'lighter' than air it can be collected by upward delivery, as shown in Fig. 4. If *dry* ammonia is required, it is passed through a drying-tube containing lumps of quick-lime.

Mix together 2 gm. of ammonium chloride and 4 gm. of slaked lime and put the mixture in a test-tube. (N.B.—The mixture must not be packed tightly in the test-tube.) Clamp the test-tube in an inclined position and heat gently. Put a few drops of concentrated hydrochloric acid in another test-tube and blow across its mouth towards the delivery-tube.

What happens when ammonia begins to escape?.....

.....

What is formed?.....

Write an equation for this reaction .....

6605

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Collect the gas in inverted test-tubes as shown in Fig. 4.

First Tube. When the tube is full of ammonia (at which point a piece of red litmus-paper held near the mouth of the tube will turn blue), close the mouth of the test-tube with the thumb and invert it in a dish of water, removing the thumb when under water.

What happens? .....
Why? .....
Test the liquid in the tube with red litmus-paper.
What has been formed? .....

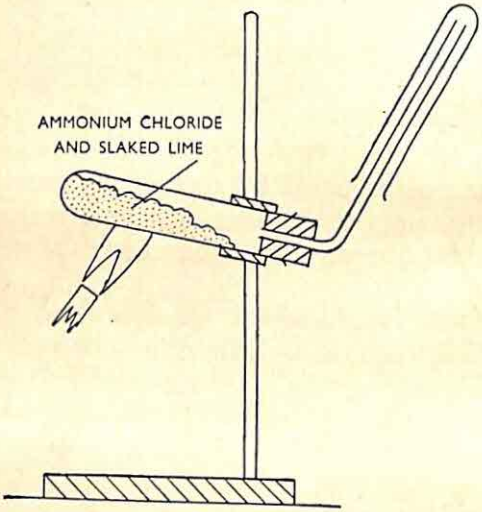


FIG. 4. Preparation of ammonia.

Second Tube. Test a tube of the gas with a lighted wooden splinter.

Does ammonia burn? .....
Does ammonia support combustion? .....

PROPERTIES OF AMMONIA

Ammonia is a colourless gas with a characteristic pungent smell. The gas is much 'lighter' than air. It is the most soluble gas known

and 1 c.cm. of water at 0° C. dissolves about 1300 c.cm. of ammonia at ordinary atmospheric pressure. Since the solubility of any gas decreases with rise of temperature, dissolved ammonia is driven off by boiling, and this is an easy way of getting a small quantity of the gas in the laboratory. At ordinary temperatures, ammonia gas can be liquefied by a pressure of 7 atmospheres (100 lb. wt. per sq. in.). When this pressure is reduced to one atmosphere, the liquid ammonia boils (at −34° C.) and as its latent heat of evaporation is 330 calories per gram, it is used in some cooling-machines (see Book Three, Chap. I).

Ammonia gas does not support combustion, and although it does not burn in air, it will burn *in oxygen*, forming nitrogen and steam:

$$4NH_3 + 3O_2 \longrightarrow 2N_2 + 6H_2O$$

This reaction shows that ammonia is a compound of nitrogen and hydrogen.

If a mixture of ammonia and air is passed over a heated catalyst, it burns to form oxides of nitrogen that dissolve in water to form nitric acid. The following equation shows the beginning and the end of a complicated series of reactions:

$$NH_3 + 2O_2 \longrightarrow HNO_3 + H_2O$$

We shall see later that this is the basis of an industrial method for making nitric acid on the large scale.

The solution of ammonia in water turns litmus blue and reacts with acids to form *ammonium salts*. Ammonia solution is therefore regarded as a *base* and is believed to contain the compound *ammonium hydroxide*, NH_4OH, in which the 'ammonium' group of atoms (NH_4) acts like a metal:

$$\underset{base}{NH_4OH} + \underset{acid}{HCl} \longrightarrow \underset{salt}{NH_4Cl} + \underset{water}{H_2O}$$

Ammonium salts can also be made by direct combination between ammonia gas and an acid, without forming water, e.g.:

$$NH_3 + HCl \longrightarrow NH_4Cl$$

Some of the ammonia of commerce is a by-product from the

manufacture of coal-gas (see pp. 59–62); but large quantities are now made from its elements (see p. 26). Ammonia is used in some types of cooling-machines, but most of the ammonia made commercially is converted into nitric acid or used in the manufacture

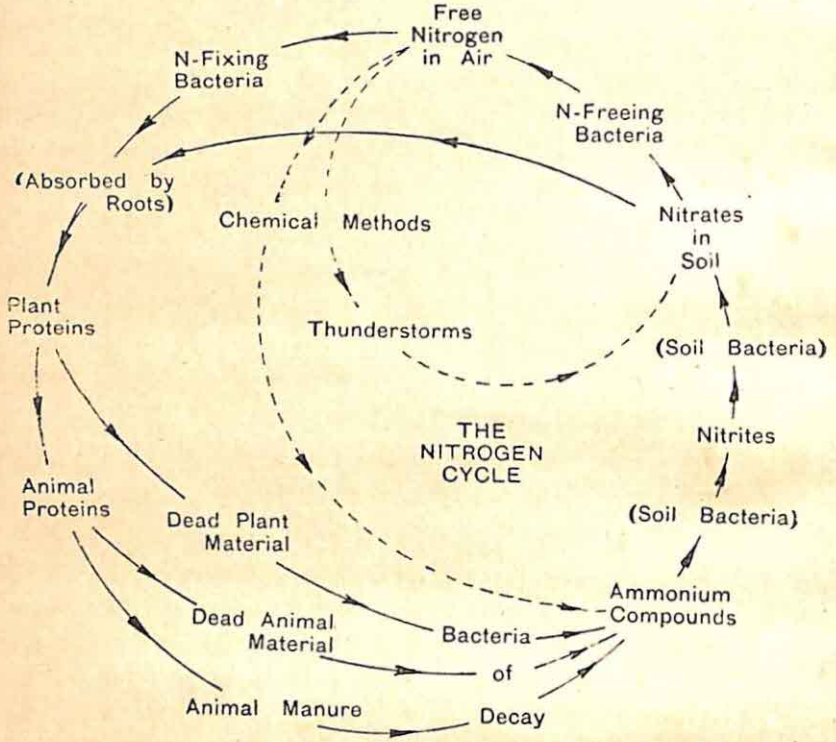


FIG. 5. The nitrogen cycle.

of explosives and fertilizers. Ammonia takes part also in the natural *nitrogen cycle*, for it is formed during the bacterial breakdown of plant and animal *proteins* in the soil (see Fig. 5).

'FIXING' ATMOSPHERIC NITROGEN

Nitrogen is an essential element for living organisms, but it is scarce in the Earth's crust, although for many years the main

supply of nitrates came from the almost rainless parts of South America, where large deposits of sodium nitrate (Chile saltpetre) are found. This supply of combined nitrogen was increased when *ammonium sulphate* was obtained as a by-product of coal-gas manufacture. But the atmosphere provides a practically inexhaustible supply of nitrogen, and methods for the 'fixation' of the element (i.e. its conversion into useful compounds) have long been sought by chemists. In Nature, nitrogen 'fixation' takes place during thunderstorms, the first reaction being the combination of oxygen and nitrogen. The energy required for this reaction comes from the lightning flash. Some of this 'fixed' nitrogen reaches the soil dissolved in rain-water (see Book Two, Chap. VI). A much more valuable natural 'fixation' of nitrogen is carried out by the *nodule bacteria*† that live in the roots of leguminous plants (see Book Two, Chap. VI). A great deal of the nitrogen 'fixed' in the bacteria passes into the host-plant and finally gets into the soil or is passed on to animals. A few years ago it was estimated that, in England alone, 600,000 tons of available nitrogen were added to the soil yearly by this means, while only 55,000 tons were added in the form of artificial fertilizers. Nevertheless, artificial nitrogen fertilizers have greatly increased farm production, especially in Europe and America. Nitrogen compounds are also used in vast quantities in making explosives (see p. 29).

Three methods have been invented by chemists for 'fixing' nitrogen:

(1) Calcium carbide (CaC_2) is heated with nitrogen in an electric furnace and forms *calcium cyanamide*,† $CaCN_2$, which can be used directly as a fertilizer. The process requires cheap electric power and is carried out, for example, at Niagara Falls.

(2) In Norway (where electricity can also be made very cheaply by water power) the natural 'fixation' of nitrogen in the lightning flash is imitated by passing air through a large electric arc (see p. 351). If the resulting gases are rapidly cooled, they contain about 1 per cent. of *nitric oxide*.† They are then passed through tall towers where oxidation takes place to *nitrogen peroxide*,† and by

absorbing this compound in water, *nitric acid* and more nitric oxide are produced. Most of the nitric acid is neutralized with limestone and sold as 'Norwegian saltpetre'—calcium nitrate.

(3) The main modern source of nitrogen compounds is *synthetic* ammonia*. This is made in the Haber Process† by direct combination of nitrogen and hydrogen. The process uses little energy and

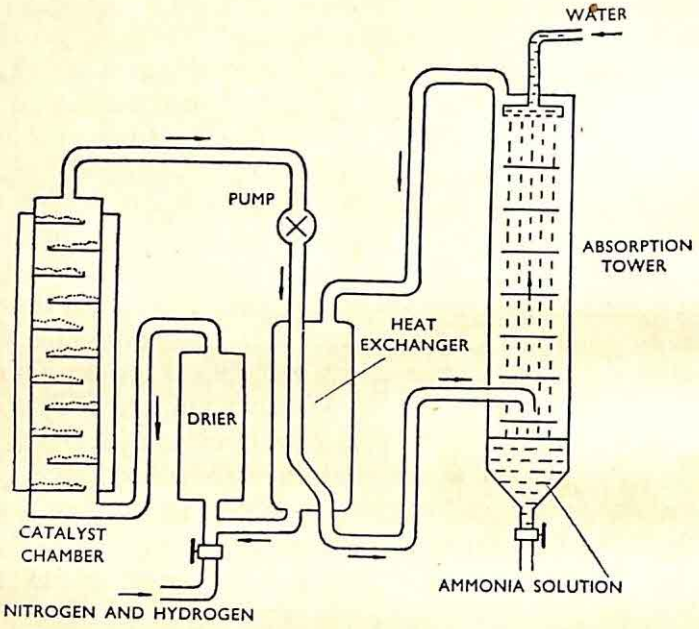


FIG. 6. Commercial synthesis of ammonia (diagrammatic).

is carried out in Europe and North America. Carefully purified nitrogen (from the air) and hydrogen (from water), under 200 atmospheres pressure, are passed over a heated catalyst (iron, mixed with small quantities of other substances) and they combine together to form ammonia (see Fig. 6).

Much of this synthetic ammonia is converted into ammonium sulphate. If nitrates are required, the ammonia can be oxidized to nitric oxide by mixing it with air and passing the mixture over a

heated catalyst. The nitric oxide is then treated as in the Norwegian process.

NITRIC ACID

Nitric acid is prepared in the laboratory (and sometimes commercially) by heating *sodium nitrate* (or potassium nitrate) with *concentrated sulphuric acid*:

$$NaNO_3 + H_2SO_4 \longrightarrow NaHSO_4 + HNO_3$$

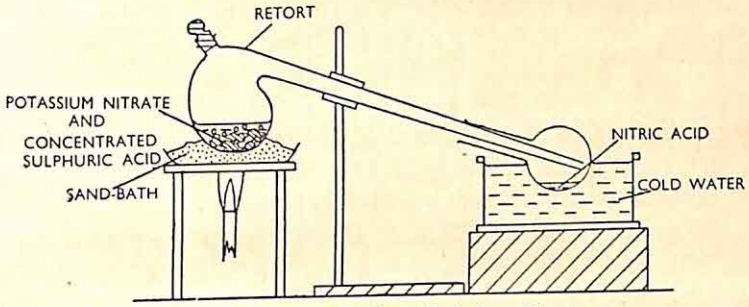


FIG. 7. Preparation of nitric acid.

At higher temperatures, sodium displaces all the hydrogen in the sulphuric acid, forming *normal sodium sulphate*:

$$2NaNO_3 + H_2SO_4 \longrightarrow Na_2SO_4 + 2HNO_3$$

but much of the nitric acid is decomposed at this higher temperature; hence only moderate heat is used and the reaction is stopped at the first stage. The sodium hydrogen sulphate is used for the manufacture of hydrochloric acid and sodium sulphate.

$$NaCl + NaHSO_4 \longrightarrow HCl + Na_2SO_4$$

Much nitric acid is now manufactured by synthetic methods, e.g. by the oxidation of synthetic ammonia.

As concentrated nitric acid is one of the most dangerous chemicals used in the laboratory, your teacher will show you how it is prepared, using the apparatus shown in Fig. 7, and putting about 15 gm. of nitrate and 25 c.cm. of concentrated sulphuric acid in the retort.

The properties of nitric acid have already been examined and described in Book Two, Chap. III

OXIDES OF NITROGEN

Nitrogen forms at least five oxides, and these five compounds of nitrogen and oxygen provide a good example of the *Law of Multiple Proportions* (see pp. 3–5). The relative proportions of nitrogen and oxygen in the various oxides are as follows:

	Nitrogen	Oxygen	
Nitrous oxide† (nitrogen monoxide)† N_2O . .	14	8	$= 1\times8$
Nitric oxide (nitrogen dioxide)† NO . . .	14	16	$= 2\times8$
Nitrous anhydride† (nitrogen trioxide)† N_2O_3 . .	14	24	$= 3\times8$
Nitrogen peroxide (nitrogen tetroxide)† N_2O_4 . .	14	32	$= 4\times8$
Nitric anhydride† (nitrogen pentoxide)† N_2O_5 . .	14	40	$= 5\times8$

Starting with the oxide with least oxygen, we find that *for every* 14 *gm. of nitrogen, the amount of oxygen increases by steps of* 8 *gm.*

Nitrous Oxide. Nitrous oxide is sometimes used by dentists,* since it produces unconsciousness and insensibility* to pain when breathed. It is prepared by heating ammonium nitrate:

$$(NH_4).NO_3 \longrightarrow 2H_2O+N_2O$$

It is a colourless gas with a sweetish smell, and, like oxygen, it makes a red-hot wooden splinter burst into flame; but, unlike oxygen, it does not form brown fumes with nitric oxide (see below).

Nitric Oxide. Nitric oxide is obtained when copper is treated with ordinary concentrated nitric acid diluted with an equal volume of water:

$$3Cu+8HNO_3 \longrightarrow 3Cu(NO_3)_2+2NO+4H_2O$$

Nitric oxide is a colourless gas that combines readily with oxygen, forming brown fumes of nitrogen peroxide:

$$2NO+O_2 \longrightarrow 2NO_2$$

Nitrogen Peroxide. Nitrogen peroxide is most easily prepared by heating lead nitrate:

$$2Pb(NO_3)_2 \longrightarrow 2PbO+4NO_2+O_2$$

It is separated from the oxygen formed at the same time by passing

the mixture through a U-tube surrounded by a freezing mixture, when the nitrogen peroxide condenses to a liquid.

EXPLOSIVES

Explosives are substances that, by chemical action, suddenly produce large quantities of heated gases that exert an enormous pressure and thus cause an *explosion*.

The oldest explosive is *gunpowder*, a mixture of *nitre* (potassium nitrate), *sulphur*, and *carbon* (charcoal). In the open it burns fiercely with a bright flash, and only explodes if it is packed tightly into a confined* space and then fired.

Gunpowder has been replaced by more modern explosives that are all *nitro-compounds* formed by the action of nitric acid on organic substances. A typical example of this type of explosive is formed when *glycerine* is treated with nitric acid, forming *nitro-glycerine*,† an oily liquid that explodes with terrible violence when given a sudden shock. It is much safer when absorbed in porous earth, and the mixture is then called *dynamite*.† *Gun-cotton*,† another very powerful explosive, is made by the action of nitric acid on cotton. *Blasting gelatine*† is a mixture of nitro-glycerine and gun-cotton. *Cordite*,† a much safer explosive used in rifle cartridges,* is a mixture of gun-cotton, nitro-glycerine, and vaseline (petroleum jelly).

The explosives used in big guns are of two different types: (1) *propellents*,* e.g. gunpowder and cordite, which burn comparatively slowly and push the shell or bullet out of the barrel without bursting the gun; and (2) *high explosives*, e.g. T.N.T. (tri-nitro-toluene),† which decompose so rapidly that they exert a shattering* effect when the shell reaches its objective. These high explosives require a small explosion to set them off; they do not explode when a light is applied to them.

SULPHUR AND ITS COMPOUNDS

Sulphur is found near the surface of the earth in regions of volcanic* activity, e.g. in Sicily, Italy, Japan, and Iceland. There is also a big deposit of sulphur in Louisiana, U.S.A., at a depth of

700 to 900 feet, but this is not thought to have been formed by volcanic action.

THE EXTRACTION OF VOLCANIC SULPHUR

In Sicily, where the natural sulphur is mixed with earth and rock, the sulphur ore is placed in kilns and melted, usually by setting

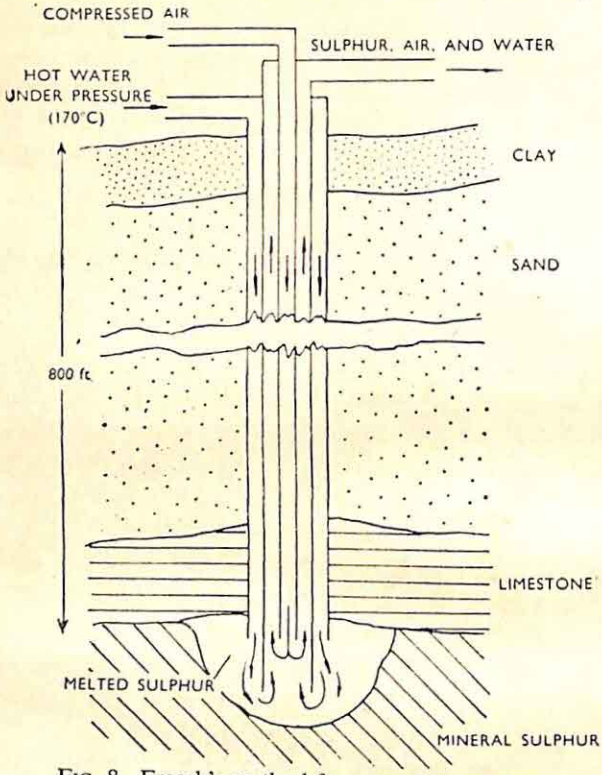


FIG. 8. Frasch's method for extracting sulphur.

fire to the sulphur. Part of the sulphur (about one-third) burns away, melting the remainder, which flows down the sloping floor into moulds, where it solidifies. Although this may seem a very wasteful method, sulphur is a cheaper fuel in Sicily than coal or wood.

This crude sulphur is further purified by *distillation*, usually in a

country where fuel is cheaper. The crude sulphur is melted and then boiled in iron retorts, from which the vapour passes into large brickwork chambers where it condenses, first to a fine pale-yellow powder, called '*flowers of sulphur*', when the chamber is cold, and later to liquid sulphur when the condensing chamber becomes hot. This liquid sulphur is run off into moulds and cast into cylindrical sticks—'*roll sulphur*'.

THE EXTRACTION OF AMERICAN SULPHUR

In Louisiana, where the sulphur is very deep down in the earth, it is extracted by the *Frasch Process*† (so called after the inventor).

Four pipes, of diameter 1 in., 3 in., 6 in., and 10 in., are sunk into the earth, one inside the other. For simplicity, Fig. 8 omits the outermost 10 in. pipe. Hot water, *superheated* to about 170° C. under great pressure, is sent down the two outside tubes to melt the sulphur. Compressed air is then blown down the central 1 in. tube, and the mixture of hot water and liquid sulphur is forced to the surface through the 3 in. pipe. This sulphur, obtained by the Frasch method, is over 99 per cent. pure and does not require any further refining.* This process supplies about three-quarters of the world's output of sulphur. Lower-grade sulphur is obtained by roasting* or distilling mineral sulphides, e.g. *iron pyrites*† (iron disulphide† FeS_2).

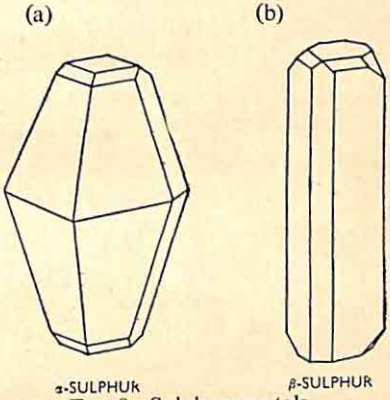


FIG. 9. Sulphur crystals.

ALLOTROPIC FORMS OF SULPHUR

The element sulphur can exist in several different forms, called *allotropes*† or *allotropic forms* of sulphur. If roll sulphur is dissolved in carbon disulphide and the solution is allowed to evaporate, crystals of *ordinary sulphur* (or *α-sulphur*) are obtained (see Fig. 9(*a*)). This is one of the allotropic forms of sulphur, and it is the

stable* form at ordinary temperatures, the other forms of sulphur turning into α-sulphur after standing for some time.

Needle-sulphur, or β-sulphur, is formed whenever liquid sulphur crystallizes on cooling. Your teacher will prepare some β-sulphur by melting roll sulphur in a large crucible and then allowing it to cool until a crust just begins to form on the surface. Two holes are pierced in this crust and the still-liquid part is then poured off. When the crust is removed, the inside of the crucible is found to be lined with transparent, needle-shaped crystals of *β-sulphur* (see Fig. 9(*b*)). On standing for a few days, the crystals become brittle, and crumble into tiny crystals of α-sulphur.

Plastic sulphur, a third allotropic form of sulphur, is made by heating sulphur to near its boiling-point (444° C.) and then pouring the hot, liquid sulphur in a thin stream into cold water. Soft elastic threads of plastic sulphur are formed, which can be moulded between the fingers. This form of sulphur is unstable* and becomes hard and brittle on standing, gradually changing back into α-sulphur.

USES OF SULPHUR

The element sulphur is of very great everyday importance. Since it burns so readily, it is used for making fireworks, gunpowder, and matches. Sulphur is used in large quantities for vulcanizing† (or hardening) rubber. Raw rubber and sulphur are heated together, and the higher the proportion of sulphur the harder is the final product. *Vulcanite*† (or *ebonite*†) is rubber that has been hardened in this way with about 30 per cent. of sulphur. Sulphur is also used in large quantities as a *fungicide*† for treating plant-diseases (and even human skin-diseases) caused by the growth of a *fungus*. Sulphur is present in the molecules of some plant and animal proteins and is thus essential to life.

COMPOUNDS OF SULPHUR

HYDROGEN SULPHIDE (OR SULPHURETTED HYDROGEN) H_2S

Sulphides, when treated with dilute acids, yield hydrogen sulphide, a poisonous gas smelling like bad eggs. For example:

$$FeS + 2HCl \longrightarrow FeCl_2 + H_2S$$

Hydrogen sulphide is also formed when proteins containing sulphur undergo bacterial decay.

In the laboratory, hydrogen sulphide is usually prepared by the action of dilute hydrochloric acid on iron sulphide, as shown by the equation on p. 32. A *Kipp's apparatus* (see Fig. 10) is generally used,

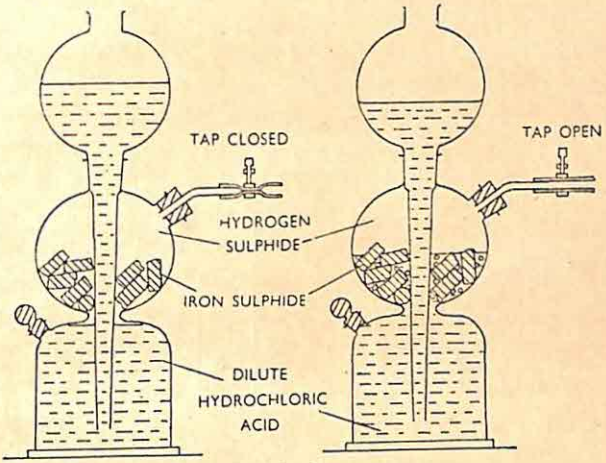


FIG. 10. Kipp's apparatus.

so that the supply of gas may be controlled. Lumps of iron sulphide are put into the middle bulb, and the acid is poured in at the top until the bottom bulb is full and the acid rises to cover the sulphide. The side tube from the middle bulb is then closed, and the pressure of the gas forces the acid away from the sulphide, thus stopping the action until the side tube is opened again.

PROPERTIES OF HYDROGEN SULPHIDE

Hydrogen sulphide is a colourless gas with a strong and unpleasant smell. It is slightly denser than air and moderately soluble in water. It causes headache when breathed in small quantities and death if breathed in large quantities. The gas burns in air with a blue flame. With a limited supply of air, e.g. when the gas is burnt in a test-tube or gas-jar, sulphur is set free:

$$2H_2S + O_2 \longrightarrow 2H_2O + 2S$$

With plenty of air or oxygen, e.g. when the gas is burnt at a jet, the sulphur also burns, forming sulphur dioxide:

$$2H_2S + 3O_2 \longrightarrow 2H_2O + 2SO_2$$

Hydrogen sulphide is an important reagent in *qualitative** *analysis*† because many metallic sulphides are insoluble and are therefore precipitated when hydrogen sulphide is passed through solutions of metallic salts.

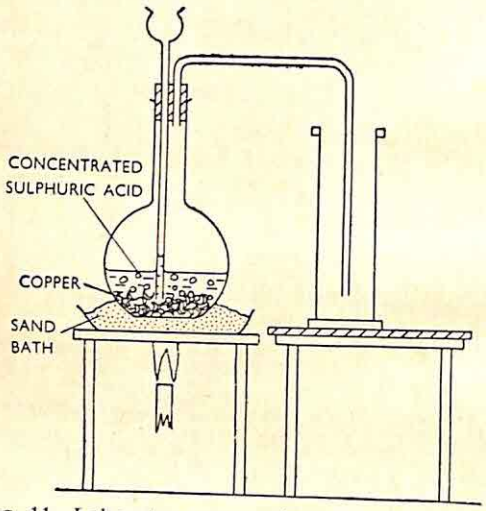


FIG. 11. Laboratory preparation of sulphur dioxide.

SULPHUR DIOXIDE SO_2

We have seen already that sulphur dioxide is formed when sulphur burns in air or oxygen:

$$S + O_2 \longrightarrow SO_2$$

In the laboratory, however, sulphur dioxide is usually prepared by the action of metals on warm, *concentrated* sulphuric acid. Copper shavings are generally used, in the apparatus shown in Fig. 11. The flask is heated and the gas is collected by downward delivery. The gas has the taste and choking* smell of burning sulphur. If a tube of the gas is inverted in water, the water rises in the tube,

showing that the gas is soluble in water. This solution turns blue litmus red, showing that an *acid* had been formed—*sulphurous acid.* A coloured flower, placed in a jar of sulphur dioxide, is bleached.

PROPERTIES OF SULPHUR DIOXIDE

Sulphur dioxide is a colourless gas with a characteristic choking smell. It is a very heavy gas: about two and a half times heavier than air. The gas is poisonous if breathed in large quantities, and it can be used as a *disinfectant** for treating rooms after cases of infectious disease. It dissolves very readily in water to form a weak acid—*sulphurous acid*:

$$H_2O + SO_2 \rightleftarrows H_2SO_3$$

When this solution is boiled, the sulphur dioxide is driven off again. Sulphurous acid slowly takes up oxygen from the air, forming *sulphuric acid*:

$$2H_2SO_3 + O_2 \longrightarrow 2H_2SO_4$$

Sulphur dioxide is readily liquefied and is sold commercially in this liquid form in strong glass bottles. Liquid sulphur dioxide (like liquid ammonia) is used in some cooling machines. The moist gas is used as a mild bleaching agent for delicate materials like silk, wool, straw (for straw hats), bristles (for toothbrushes), and sponges, which would be damaged if bleached by chlorine. Chlorine bleaches by *oxidation* and the coloured molecules are decomposed, while sulphur dioxide bleaches by *reduction.* Hence materials bleached by sulphur dioxide gradually regain their original colour when exposed to the air, e.g. newspapers turn yellow in time because the paper is made from wood-pulp that has been bleached with sulphur dioxide. Sulphur dioxide does not burn itself, and will not allow a lighted candle to burn in it, but a few substances like magnesium, which burn very vigorously, continue to burn in the gas, setting free sulphur:

$$2Mg + SO_2 \longrightarrow 2MgO + S$$

This is a simple way of proving that sulphur dioxide contains

sulphur. It should remind you of a similar experiment, burning magnesium in carbon dioxide:

$$2Mg + CO_2 \longrightarrow 2MgO + C$$

SULPHURIC ACID

SULPHUR TRIOXIDE, SO_3

If a mixture of dry sulphur dioxide and oxygen is passed over finely divided platinum in a hot tube, *sulphur trioxide* is formed, the platinum acting as a catalyst:[1]

$$2SO_2 + O_2 \longrightarrow 2SO_3$$

Sulphur trioxide is a white solid that crystallizes in long, silky needles. It has a great liking for water, dissolving with a hissing noise to form *sulphuric acid*:

$$H_2O + SO_3 \longrightarrow H_2SO_4$$

Hence, to make sulphuric acid, the most direct way is to prepare sulphur trioxide and then dissolve it in water.

THE MANUFACTURE OF SULPHURIC ACID

Sulphuric acid is the most important of all commercial chemicals, and millions of tons of it are used each year. It is manufactured commercially by two main methods:

(i) the older *lead-chamber process*; (ii) the more modern *contact process*.

THE LEAD-CHAMBER PROCESS

In this process, jets of *water* or *steam* are blown into large leaden chambers along with a mixture of *sulphur dioxide* and *air* containing also *oxides of nitrogen* that act as a catalyst in the oxidation of sulphur *di*-oxide to sulphur *tri*-oxide. The following equations represent the main changes that take place:

$$H_2O + SO_2 + NO_2 \longrightarrow H_2SO_4 + NO$$
$$NO + \text{oxygen from the air} \longrightarrow NO_2$$

[1] As the gaseous reaction takes place at the *surface* of the solid catalyst, it is important to use the catalyst in a finely divided form that provides the maximum surface area for the minimum volume of solid catalyst.

Nitric oxide (NO) serves as an '*oxygen-carrier*', first taking up oxygen from the air to form *nitrogen peroxide* (NO_2) and then giving it up again to the sulphur dioxide.

The acid that forms in the lead chambers is only of about 65 per cent. strength and, after concentrating by evaporation, it is sold as 'commercial sulphuric acid'. The chamber acid is relatively cheap, but it is also impure; hence it is used for making 'heavy chemicals' and fertilizers for agricultural purposes.

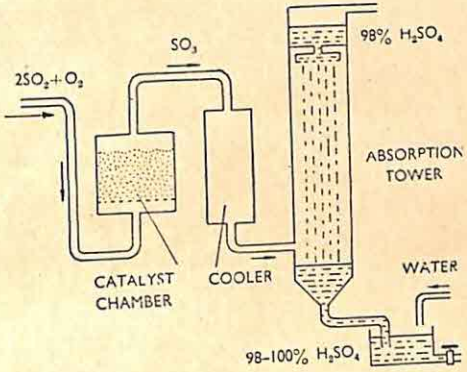


FIG. 12. Contact process (diagrammatic).

THE CONTACT PROCESS

In the lead-chamber process nitric oxide is used as the catalyst for making the sulphur dioxide combine with oxygen to form sulphur trioxide. The contact process uses finely divided vanadium pentoxide† as a catalyst instead of nitric oxide, and it is essentially the same method as that described on p. 36 for making sulphur trioxide.

Dry sulphur dioxide and air are passed over the heated catalyst in a '*catalyst chamber*' and *sulphur trioxide* is formed (see Fig. 12).

$$2SO_2+O_2 = 2SO_3$$

The sulphur trioxide formed is absorbed in 98 per cent. sulphuric acid, since unpleasant fumes are formed when water is used for dissolving the sulphur trioxide. Water is added gradually to the

acid, combining with the sulphur trioxide as it dissolves and maintaining a 98 per cent. concentration of acid.

Sulphuric acid prepared by the contact process is much purer than that from the lead-chamber process, although it is somewhat more expensive. If the price of the sulphuric acid from the contact process could be reduced, the lead-chamber process could not compete with it. Where a concentrated acid is required the contact process is the best source; but for a more dilute acid the chamber process is satisfactory and economical.

The purer acid is used for lead accumulators (batteries), for making 'fine chemicals' and dyestuffs, and in some processes for refining foodstuffs.

We have already examined the properties of sulphuric acid (Book Two, Chap. III).

ALKALI MANUFACTURE

The manufacture of *alkalis* is a close second in importance to the manufacture of sulphuric acid. The most important commercial alkalis are *washing soda* (sodium carbonate), *caustic soda* (sodium hydroxide), and *caustic potash* (potassium hydroxide).

Until the end of the eighteenth century, the only source of alkali for making soap and glass was *plant ashes*. In 1791, Leblanc, a French scientist, invented the first successful method for making washing soda from common salt. Leblanc's process has now been replaced by the *ammonia-soda process*.

SOLVAY'S PROCESS (THE AMMONIA-SODA PROCESS)

This is a more modern and a simpler process than Leblanc's method. *Brine* is saturated with *ammonia* and then *carbon dioxide* is passed in. *Ammonium bicarbonate*† is first formed from the water, ammonia, and carbon dioxide:

$$H_2O + CO_2 + NH_3 \longrightarrow (NH_4)HCO_3$$

This ammonium bicarbonate then reacts with the sodium chloride in the brine to form *sodium bicarbonate*† (baking soda†), which,

being less soluble in water than the other substances, settles out from the liquid:

$$NaCl + (NH_4)HCO_3 \longrightarrow (NH_4)Cl + NaHCO_3$$

This sodium bicarbonate is easily converted into *anhydrous sodium carbonate*† (soda-ash) by heating it:

$$2NaHCO_3 \longrightarrow Na_2CO_3 + CO_2 + H_2O$$

Millions of tons of soda-ash are used every year in the manufacture of glass,* soap, textiles,* chemicals, and paper.

CAUSTIC ALKALIS (CAUSTIC SODA AND CAUSTIC POTASH)

There are two commercial methods for manufacturing caustic soda; (i) by the action of lime on sodium carbonate, (ii) by the electrolysis of brine. The first method supplies the greater part of present-day production.

(i) A 10 per cent. solution of sodium carbonate is boiled in a large iron vessel and quicklime is placed in an iron cage dipping into the solution. Most of the sodium carbonate is converted into sodium hydroxide and calcium carbonate is precipitated:

$$Na_2CO_3 + Ca(OH)_2 \longrightarrow 2NaOH + CaCO_3$$

The solution of sodium hydroxide is filtered off and concentrated by evaporation under reduced pressure. This concentrated solution is finally evaporated to dryness in cast-iron vessels and fused sodium hydroxide is left.

(ii) At first sight, it seems an easy matter to obtain *sodium hydroxide* (caustic soda) by electrolyzing a solution of sodium chloride. One would expect metallic sodium to be set free at the cathode† (negative electrode) and to react at once with the water to form caustic soda, while chlorine should be set free at the anode† (positive electrode) and should escape. Unfortunately, some of this chlorine combines with the caustic soda unless special precautions are taken to keep the chlorine away from the alkali. There are several methods for doing this on the large scale, but we shall consider only one of these—the *Castner–Kellner process*.†

In this process saturated brine runs slowly into one end of a rectangular tank (see Fig. 13). The floor of the tank slopes slightly towards the other end and a slow stream of mercury flows down this slope. This mercury forms the *cathode* (negative) and carbon *anodes* (positive) hang in the brine. When an electric current is passed, chlorine is set free at the anodes and is led away through a pipe in the roof of the tank. Sodium is set free at the surface of

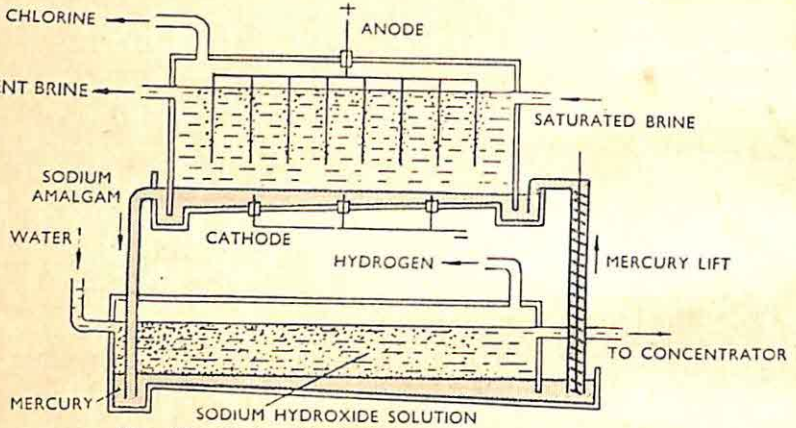


FIG. 13. Castner-Kellner process (diagrammatic).

the mercury cathode and *dissolves in the mercury*. This solution (*sodium amalgam*†) is run off into a lower tank containing water, where the sodium reacts to form sodium hydroxide and hydrogen. This sodium hydroxide is then evaporated. The chlorine that escapes is a valuable by-product and is often used for making *bleaching-powder* (Book Two, Chap. II). Sometimes, however, the hydrogen that escapes from the lower chamber is burnt in the chlorine to form hydrogen chloride, which is dissolved in distilled water and sold as chemically pure hydrochloric acid.

METALS AND NON-METALS

It is convenient to divide the elements into two classes—*metals* and *non-metals*. The most important *chemical* property that distinguishes these classes of elements is that *metals form basic*

oxides, while non-metals do not; *most non-metals form acidic oxides*. The most important *physical* property of a metal is its *toughness*. Metals are strong and tough, not brittle (like glass), and we use them for many everyday purposes because they can resist a strong pulling force without breaking. Connected with this toughness is their capacity to be forged,* rolled into very thin sheets, pressed into different shapes, and spun or drawn into fine wire. Most metals have a high density and a high melting-point (in contrast to most non-metals) and all except mercury are solid at ordinary temperatures. Unlike non-metals, most metals can be polished. They also differ from non-metals in being good conductors of heat and of electricity, properties that add to their everyday usefulness.

All metals and many non-metals are *crystalline*; this is not obvious in the external form of many metals, the crystals being microscopic, but many of the properties of metals are associated with their crystalline structure.

METALS IN EVERYDAY LIFE—CORROSION

For a metal to be useful for everyday purposes, it must resist the action of air and water. The very active metals *potassium*, *sodium*, and *calcium*, which react with cold water, can only be used for chemical reactions in laboratories. *Magnesium*, also, is too active for most everyday purposes, but since it is one of the 'lightest' metals (S.G. 1·7) it is mixed with aluminium and other metals for making aeroplane parts. Metals in general use that do not rust in air are *aluminium*, *copper*, *nickel*, *tin*, *zinc*, *lead*, *silver*, *gold*, and *mercury*. Gold, silver, and nickel are quite unaffected by air or water. Aluminium, zinc, and lead rust only on the surface, for the film of oxide clings to the surface and so protects the metal underneath from any further action. (This also happens to a slight extent with magnesium, as you have noticed when using magnesium ribbon.)

Iron is the cheapest commercial metal but it has the great disadvantage that it corrodes or rusts quickly in damp air, and that

the rust does not cling to the surface but falls off, exposing a fresh surface of iron until the whole of the iron is converted into rust. Iron articles, therefore, must be protected from the action of air in some way, e.g. by painting, by coating with tin (tin-plate), or zinc (galvanized iron), or silver (silver-plate), or nickel (nickel-plate), or chromium (chromium-plate).

ALLOYS†

'Chemically pure' metals are not used for industrial purposes. 'Commercially pure' metals, containing over 99 per cent. of one metallic element, are sometimes used for special purposes, but usually another element (or more than one) is added to the melted metal to form an *alloy*. This is done to get special properties, e.g. *brass*, an alloy of copper and zinc, combines the most useful properties of both metals. A very small proportion of the alloying element is often sufficient to produce very marked changes in properties. This alloying element is usually another metal but is sometimes a non-metal, e.g. carbon in steel. Alloys are generally harder, less tough, and with lower electrical and heat conductivities than the pure metals. Thus, almost pure copper (99·98 per cent.) is used for electrical wiring but is too soft for coinage. An alloy containing 95 per cent. of copper and 5 per cent. of tin or zinc is used for making 'copper' coins because this alloy is much harder than pure copper. *Gun-metal*† and *bronze*† are alloys of copper with tin. *Pewter*† and *solder* are alloys of tin and lead. '*German silver*'† is an alloy of copper, zinc, and nickel. *Type-metal*, for making printers' type, is an alloy of lead, tin, and antimony. *Bearing metal* (or *anti-friction metal*), for machine bearings, consists of the same three metals mixed in different proportions.

Many '*light alloys*' are used for purposes where weight must be kept as low as possible, e.g. in aeroplane parts and the crank-cases of motor-engines. '*Magnalium*'† and '*Duralumin*'† are mainly aluminium (S.G. 2·7) with small quantities of magnesium, copper, nickel, and manganese. '*Elektron*'† (90 per cent. mag-

nesium with some copper, zinc, manganese, and silicon) is one of the 'lightest' of these alloys, with a specific gravity of only 1·8.

Other alloys we have met with in science lessons are '*Invar*' (65 per cent. iron, 35 per cent. nickel), and '*Nichrome*' (60 per cent. nickel, 15 per cent. chromium, and 25 per cent. iron).

MINERAL ORES

Only a few chemically inactive metals, e.g. gold and platinum, are found *native* in the Earth's crust. Most metallic elements are found combined with non-metallic elements as *minerals*. From some of these—the *ore minerals*—it is fairly easy to extract the metal. Thus, aluminium-containing *minerals* are very common in the Earth's crust, in clays and in most rocks, but the metal can, at present, only be extracted commercially from the *ore*† mineral bauxite,† which is comparatively scarce. The ores of iron and tin are mainly oxides; those of copper, zinc, lead, and mercury are sulphides. The extraction of metals from ores is a branch of *metallurgy*,† and the extraction processes, which vary with different ores, are usually complicated. Here in this Course, therefore, we shall deal only with the general principles of extracting metals from their ores.

GENERAL PRINCIPLES OF EXTRACTING METALS FROM ORES

The ore is first got from the earth by *mining*. It is then '*dressed*'. *Ore-dressing*† means preliminary* treatment by *crushing*, *grinding*, and *concentration* so as to liberate and separate the valuable minerals from others. The valuable ore may be concentrated by magnetic or gravity separation or by *flotation*.† In the latter process the ore particles are stirred up in water with a chemical compound that sticks to one of the mineral constituents and makes it float free of the others. Ore-dressing produces no chemical change and the dressed ore is afterwards treated in one of the following ways so as to extract the metal from it:

(*a*) *By solution* (or *leaching*).† The ore is dissolved in an aqueous* solution and the metal is later recovered from this solution, e.g.

gold can be extracted from low-grade ore by dissolving out with sodium cyanide† solution.

(*b*) *By electrolysis* of a solution of the ore, e.g. copper is purified by electrolysis of copper sulphate solution (see pp. 354-6).

(*c*) *By heat.* Ores that contain the metal combined as a carbonate are first heated strongly to drive off carbon dioxide and convert them into oxides. Sulphide ores, too, are first heated strongly in a current of air to convert the sulphur into sulphur dioxide and the metal into its oxide.

In nearly all cases, when the oxide has been obtained, this is reduced to the metal by heating with carbon. (You can do the same thing on a small scale by heating lead oxide with a blowpipe on a charcoal block.) Metallic ores are nearly always mixed with earthy matter, and during the reduction (or *smelting*†) with carbon, it is usual to add limestone or chalk. The quick-lime that is formed combines with the earthy matter to form a liquid *slag*† that floats on the top of the liquid metal and can thus be separated from it.

IRON AND STEEL

Iron is the most important and useful of all the metals. It is found mainly as iron oxide and iron carbonate ores.

THE METALLURGY OF IRON

The ore is first heated strongly to drive off carbon dioxide and water, leaving ferric oxide†—Fe_2O_3—and this oxide is then *smelted* with carbon (coke†) in a *blast* furnace.* Limestone is added to convert the earthy impurities into a liquid *slag.*

A *blast furnace* is a tall iron tower, 50–100 ft. high, lined with fire-brick. A mixture of iron ore, coke, and limestone is added at the top, and at the bottom is the *hearth** in which the liquid iron and slag collect (see Fig. 14). Just above the hearth is a ring of jets through which a blast of hot air is blown into the furnace. The coke burns to form *carbon monoxide*, which is the *reducing agent*, and the ferric oxide is reduced to metallic iron:

$$Fe_2O_3 + 3CO \longrightarrow 3CO_2 + 2Fe$$

The liquid iron and slag collect in the hearth at the base of the blast furnace, the iron below the slag, whence it is run off from time to time through the lower hole, while the slag is run off through the upper hole. In this way the blast furnace can be worked continuously.

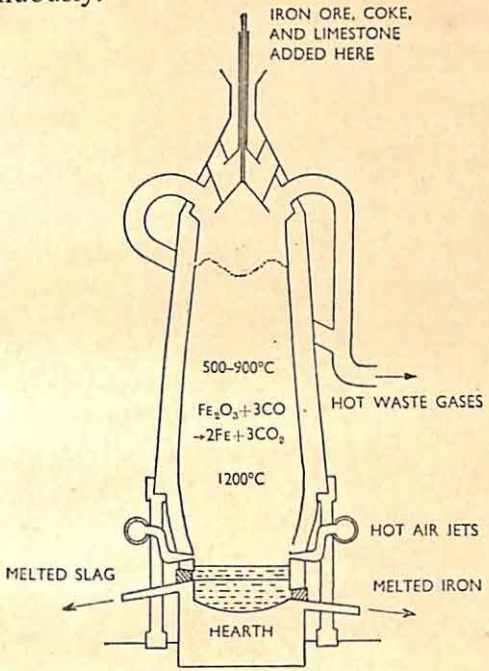


FIG. 14. Blast furnace (diagrammatic).

CAST-IRON

The liquid iron from the blast furnace is run into long, narrow moulds where it solidifies, forming '*pig-iron*'. This form of iron contains not more than 92 per cent. of iron, the remainder being mainly carbon and silicon. This impure 'pig-iron' is usually remelted with coke and limestone in a smaller furnace, and this molten iron is then '*cast*' into heavy articles that do not have to stand a bending force or a sudden shock, e.g. the *bases* of laboratory retort stands are made of *cast-iron*. In making castings, a mould of

the required shape is made in sand and liquid iron is poured into the mould, taking the shape of the containing vessel. Owing to its impurities, cast-iron is hard but brittle, breaking readily when hammered or bent.

WROUGHT-IRON†

Cast-iron melts at 1,100–1,150° C. and if this molten iron is exposed to air some of the impurities are oxidized and the melting-point rises. *Wrought-iron* is made in a furnace lined with iron oxide in which the temperature does not rise high enough to melt the product in the later stages, i.e. the molten cast-iron becomes plastic and the impurities (or slag) cannot escape from it. When this iron cools it develops a *fibrous structure* (partly due to threads of slag) and it cannot be used for making castings. Wrought-iron is tough and 'soft'. (The *rods* of laboratory retort stands are made of wrought-iron.) It can be drawn out into wire (i.e. wrought-iron is *ductile*†) or hammered into thin sheets (i.e. wrought-iron is *malleable*†). Wrought-iron softens before it melts and can therefore be *forged* (shaped) and *welded* (united) by hammering when red-hot.

Wrought-iron is the purest commercial form of iron, containing less than 0·5 per cent. of impurities (of which less than half is carbon). Its special properties, however, are due to its method of preparation rather than to its purity.

STEEL

Most kinds of steel are made by oxidizing away impurities from cast-iron. Steel is made, however, in a furnace producing a higher temperature than that used for making wrought-iron, so that the cast-iron, instead of remaining plastic, becomes liquid and the slag rises to the surface. After this the steel is cast. Steel usually contains from 0·1 to 1·5 per cent. of carbon, the main production being in the range 0·1–0·4 per cent. of carbon.

Small quantities of the best steel (which is used for making razors and cutting tools) are made by the *crucible process*, in which carbon is added to wrought-iron. This 'crucible steel' is expensive.

Ordinary cheap steel is made either by the *Bessemer process* or by the *open-hearth process*, in both of which some of the carbon is removed from cast-iron. The world's yearly production of steel by these processes is over 200,000,000 tons.

STEEL MANUFACTURE

(i) *The Bessemer Process.* In this process excess of carbon is removed from liquid cast-iron by a blast of hot air blown through it in a *Bessemer converter* (see Fig. 15). This is a large iron vessel with a basic fire-proof lining. The converter is half filled with several tons of molten cast-iron, and hot air is blown through it until most of the impurities in the iron have burned away. The necessary amount of carbon is then added to the liquid metal, the converter is turned on horizontal pivots, and the steel is poured out into moulds.

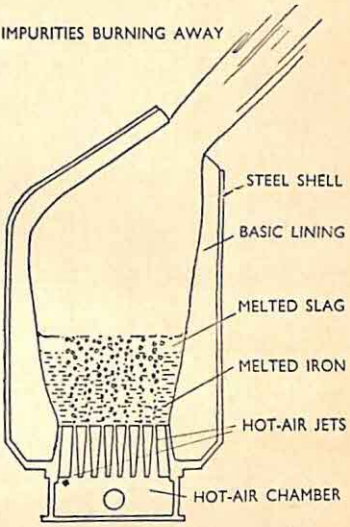


FIG. 15. Bessemer converter (diagrammatic).

The characteristic feature* of the Bessemer process is its speed. In a modern converter, 10 tons of cast-iron can be changed into steel in 10 minutes. No external fuel is needed—this is supplied by the impurities in the cast-iron. The temperature has to reach 1,500° C., and unless oxidation goes on very rapidly, too much heat is lost in the waste gases and by radiation. Although good steel can be made in this way by experienced workmen, chemical control of the process (i.e. exact regulation of the carbon content) is impossible because of the speed of the reaction. The process has now been largely replaced in England by the 'open-hearth' method. But Bessemer's invention (1856) made cheap steel available for the first time and was largely responsible for the rapid growth of industry during the last century.

(ii) *The Open-hearth Process.* In this process a mixture of liquid cast-iron, iron ore, and scrap steel is heated in an open-hearth furnace (see Fig. 16). An external fuel is used, usually producer-gas (see pp. 63–65), which is pre-heated by the waste gases from the furnace. The impurities are oxidized at this high temperature by the iron oxide that dissolves in the slag. The process need not be

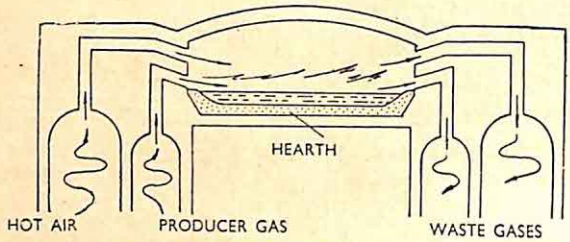


FIG. 16. Reversible open-hearth furnace (diagrammatic).

rapid (it usually takes about 10 hours) and is therefore more easily controlled than the Bessemer process, so that a steel of definite composition can be produced.

THE PROPERTIES OF STEEL

The properties of steel depend upon two main factors:

(*a*) *the percentage of carbon in the steel*;
(*b*) *the heat treatment given to the cast steel.*

The carbon content of steel varies from 0·05 per cent. to 2·0 per cent. Steel with less than 0·3 per cent. of carbon—'*mild*' *steel*—is comparatively soft, and hardness increases with increase of carbon. *Structural steel* (for bridges and steel-framed buildings) is fairly 'soft'; steel for machine parts contains from 0·2 to 0·6 per cent. of carbon and is harder. *Tool steel*, or high-carbon steel, contains from 1·0 to 1·5 per cent. of carbon and is very hard indeed. Especially hardened steel alloys containing small quantities of manganese and chromium have now been developed.

If red-hot steel is '*quenched*'† (i.e. rapidly cooled in oil or water), it becomes very hard and very brittle. This steel can be *tempered*†

by careful reheating, when it becomes tougher and less hard. By such heat treatment steel may be brought to any desired degree of hardness and toughness.

SILICON COMPOUNDS IN EVERYDAY LIFE

Next to oxygen, *silicon* is the commonest element in the Earth's crust, and its compounds are of great economic importance. Silicon dioxide (or *silica*, SiO_2) exists in Nature as *sand*, *flint*,† and *quartz*. Clay is mainly aluminium silicate. *Glass*, *pottery*, *bricks*, and *cement* are also composed mainly of silicon compounds.

Quartz, a pure form of silica, is so hard that it scratches glass and cannot be scratched by steel. It has a very high melting-point (1,500° C.), but it can be melted in the oxy-hydrogen flame and is used for making special laboratory apparatus.

Owing to its very low coefficient of expansion (0·5 per million), this quartz apparatus can be heated and cooled very rapidly without cracking.

If silica (e.g. quartz sand) is melted with sodium carbonate, it forms *sodium silicate*,† Na_2SiO_3, a concentrated solution of which is called '*water-glass*'.† This solution can be used for preserving eggs, for hardening concrete floors, for waterproofing cloth, for making asbestos sheets, and for sticking things together.

GLASS

When sand is melted with soda-ash (anhydrous sodium carbonate), 'water-glass' is formed. If sand is melted with soda-ash and also lime, ordinary bottle-glass or window-glass is produced. Glass is a mixture of silicates, and common glass is made by melting a mixture of quartz sand, limestone, and soda-ash. *Hard glass*, or potash glass, to stand higher temperatures without melting, is made in the same way, except that potash is substituted for the soda-ash. *Flint glass*, or *lead glass*, a sparkling glass of high refractive index, used for making lenses and 'cut-glass' articles, is made from sand, limestone, soda-ash, and lead oxide. If the silica in a glass is partly replaced by the oxides of *boron*† and *alu-*

minium, a 'boro-silicate glass' is obtained that will stand sudden changes of temperature without cracking. Such glass can therefore be made thicker and stronger than ordinary glassware for heating. Coloured glass is made by dissolving small amounts of various metallic oxides in liquid glass, e.g. a trace of cobalt oxide gives a blue glass.

THE PROPERTIES OF GLASS

The most important properties of glass are (*a*) that it is transparent, (*b*) that it resists the action of most chemicals, and (*c*) that it softens gradually before it melts, so that it can be worked while it is in this plastic condition. Most other solids melt sharply, without passing through a plastic or semi-solid stage. Glass resists the action of nearly all chemicals, the most important exception being *hydrogen fluoride*† (or hydrofluoric acid†), which dissolves glass, and can therefore be used for marking glass, e.g. in graduating glass apparatus. The peculiar properties of glass make it a material of the very greatest importance in modern life; in particular, we should find it very difficult to equip a scientific laboratory if no glass was available.

CLAY, BRICKS, AND POTTERY

Clay, as we learnt in Book Two, Chap. VI, is a product of the weathering of rocks containing felspar and mica. All clays contain aluminium silicates, together with incompletely decomposed mineral particles. Bricks are made by baking clay to a red heat. At this high temperature some of the silicates melt and bind the other particles into a hard mass. The effect of heating is irreversible, so that wet brick-dust no longer shows the characteristic plastic properties of unbaked clay. Pottery, china, and porcelain are made in a similar way—by 'firing' to a high temperature—but the clay is carefully chosen and is first washed to remove coarse particles. The baked product is still porous and requires *glazing*. It is coated with a paste of silica and felspar (a potassium-aluminium silicate) and is then 'fired' again in the kiln, producing a

thin surface film of glass. Coloured patterns are produced by painting the design in metallic oxides before glazing.

CEMENT

The Romans found that if a mixture of limestone and clay is heated strongly, the product forms a mortar that will harden even when it is covered with water. Many Roman buildings in which this 'hydraulic mortar' was used are still standing to this day. Such a mortar, which will harden in water as well as in air, is called *hydraulic cement*.

The most important kind of cement is called 'Portland cement', because the earliest cements of this type were said to yield concrete 'equal to the best Portland stone' for building purposes. This cement is made by strongly heating a powdered mixture of limestone and clay to a temperature of about 1,500° C. in a sloping, revolving furnace. The mixture partly melts and forms a hard mass consisting of a mixture of quicklime, calcium silicate,† and calcium aluminate.† This is ground up into a fine powder and then packed in air-tight containers.

The importance of cement increases steadily year by year. Cement mortar is used widely in place of lime mortar for building with brick and stone. When cement, mixed with sand and gravel, is made into a paste with water, it forms '*concrete*', which sets to a solid mass in a short time. Concrete is used in ever-increasing quantities for building, and also for making articles that were formerly made of wood, iron, or stone. Steel rods and wires are often embedded in the concrete to give it greater strength, forming '*reinforced concrete*'.†

Concrete owes its usefulness to the fact that it can be shaped while in the semi-liquid, plastic state. We have learnt that the most characteristic property of *solids* is *rigidity*, and that it is always a matter of some difficulty to change the shape of a solid. A *liquid*, on the other hand, is not rigid, and takes the shape of any containing vessel. Liquid concrete is poured into a temporary vessel of the desired shape (a mould), and it takes the shape of this

'containing vessel', still retaining that shape after it sets to a rigid solid.

CARBORUNDUM

When a mixture of sand and coke is strongly heated in an electric furnace, *silicon carbide*† (SiC) is formed. This substance (commonly called *carborundum*†) is extremely hard and is used for sharpening steel cutting tools.

PHOSPHORUS AND SOME OF ITS COMPOUNDS

The element phosphorus is widely distributed in small amounts. In North America, North Africa, and some Pacific islands, however, there are thick deposits of rock phosphate (*calcium phosphate*) (see also p. 171). These rock phosphates supply most of the calcium phosphate used as a fertilizer in agriculture.

Phosphorus compounds are present in all living matter. Plants take in *phosphates* from the soil, and animals obtain their phosphorus compounds from certain *proteins* synthesized by plants.

The element *phosphorus* was first prepared from animal *bones*, which contain about 50 per cent. of *calcium phosphate*. When bone charcoal is heated in air, all the carbon burns away, leaving *bone-ash*, which is mainly calcium phosphate. When this bone-ash is treated with sulphuric acid, *phosphoric acid*, a thick, sticky liquid, is obtained, and if this liquid is mixed with charcoal and then heated to a white heat in the absence of air, phosphorus distils off as a vapour and can be condensed under water. A more modern method is to heat calcium phosphate rock with quartz and coke, in an electric furnace. The silica of the quartz forms calcium silicate, displacing *phosphorus pentoxide*, which is then reduced to phosphorus by the carbon.

Phosphorus, like sulphur and carbon, exists in *allotropic forms*, the two commonest being *yellow phosphorus* and *red phosphorus*.

YELLOW PHOSPHORUS

Yellow phosphorus (sometimes called white phosphorus) is a translucent solid like bees' wax in appearance. It is white when

freshly prepared, but it gradually turns yellow on exposure to light. Yellow phosphorus has a very low *ignition temperature*, and will ignite* in air at about 30° C. (=86°F.). Hence it must always be handled with the greatest care, particularly in tropical countries. In the laboratory, yellow phosphorus is kept *under water*. This form of phosphorus is a very active *poison*, 0·1 gm. causing death. It always gives off a poisonous vapour, and if small quantities of the vapour are breathed day after day, it causes a disease in which the bones decay, beginning with the jaw-bones. At one time, this disease was very common among workers in match factories, and in most countries the use of yellow phosphorus for making matches is now forbidden by law.

RED PHOSPHORUS

When yellow phosphorus is heated to about 250° C. in the absence of air, it changes to a dark red powder—*red phosphorus*—which is another allotropic form of the element. This variety of phosphorus gives off no poisonous vapour, and it does not ignite below 240° C.; in fact, it is a much less active form.

The chemical identity* of these two forms of phosphorus can be proved by burning equal weights of yellow and red phosphorus in air or oxygen, when the same weight of the same product (phosphorus pentoxide) is formed in both cases.

MATCHES

For thousands of years, Man could only produce fire by friction between pieces of wood, or by using 'flint and steel'. Matches have only been in use for the last hundred years or so.

Matches were formerly made by dipping splinters of wood into a paste of yellow phosphorus, lead peroxide, and glue, but when the use of yellow phosphorus was stopped, a sulphide of phosphorus was substituted. Such matches will 'strike' on any rough surface, the friction producing sufficient heat to ignite the mixture.

Modern '*safety matches*' contain no phosphorus in the match-head, but a strip of paper coated with non-poisonous *red phosphorus* is attached to the side of the match-box. The match-head

contains an oxidizing agent (e.g. potassium chlorate, potassium nitrate, or red lead) mixed with glue. When the match-head is rubbed on the side of the box, the friction ignites a little of the red phosphorus and thus sets fire to the inflammable match-head, which, in turn, sets fire to the wood of the match-stick.

PHOSPHORUS PENTOXIDE (P_2O_5) AND PHOSPHORIC ACID (H_3PO_4)

When phosphorus burns in dry air or oxygen, a cloud of *phosphorus pentoxide* is formed. On standing, this settles as a white powder. Phosphorus pentoxide has a great chemical attraction for water, dissolving with much heat to form *phosphoric acid.* Because of this great attraction for water, phosphorus pentoxide is used for drying gases when every trace of water vapour must be removed, and it is the most powerful drying agent known.

Phosphates are the salts of phosphoric acid and we have seen that they are valuable fertilizers, e.g. calcium phosphate, ammonium phosphate. (See *General Science Workbook*, Book Two, pp. 135–44, for Learning Exercises on this chapter.)

Chapter III

CARBON AND SOME OF ITS COMPOUNDS

Carbon forms more compounds than any other element; in fact, the study of carbon compounds forms a special branch of chemistry—*organic chemistry*. Although the term '*organic*' suggests that these compounds are obtained from living things, most of them can now be prepared in the laboratory and some of them do not exist in Nature.

The element carbon, like sulphur, exists in *allotropic forms*: (i) *non-crystalline carbon* (e.g. sugar-charcoal), and (ii) *crystalline carbon*, namely *graphite*† and *diamond*. It is surprising to find that diamond, graphite, and charcoal should all be the same chemically, but this is proved to be so by burning the same weight of the three substances in oxygen, when equal weights of the same product, carbon dioxide, are obtained. Hence, although they are so very different in appearance, diamond, graphite, and charcoal are only *different physical forms of the same chemical element—carbon*.

There are several common kinds of impure non-crystalline carbon: (*a*) *wood-charcoal*, (*b*) *lamp-black*† or *soot*, (*c*) *animal charcoal* (or *bone charcoal*), and (*d*) *coke*.

CHARCOAL

Wood-charcoal is prepared by the *destructive distillation of wood* (i.e. decomposition without flame), either by the old method of burning wood in heaps covered with earth (see Fig. 17), or by the modern method of heating wood in closed ovens or retorts (see Fig. 18). When the latter method is used, valuable by-products are obtained, e.g. *acetic acid*,† *methyl alcohol*,† *acetone*,† *wood-tar*,† and *wood-gas*† (mainly methane). When the old-fashioned method is used, these substances escape into the air along with the smoke.

Wood-charcoal is a black, porous substance of very low density,

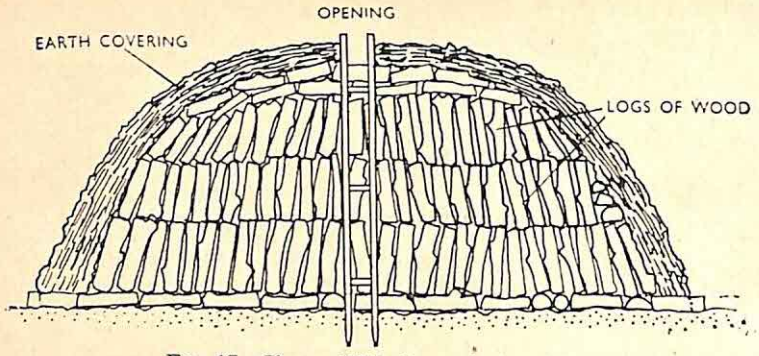


FIG. 17. Charcoal kiln in vertical section.

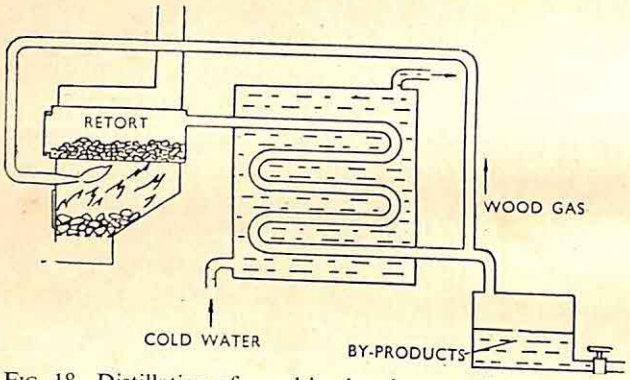


FIG. 18. Distillation of wood in closed retort (diagrammatic).

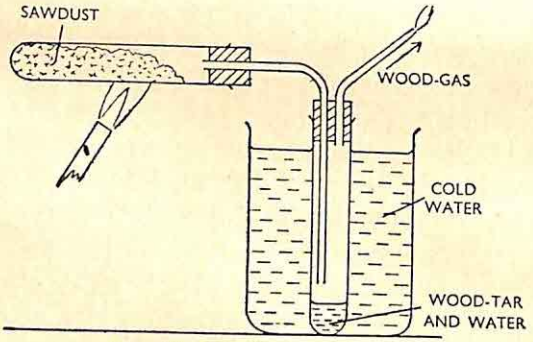


FIG. 19. Distillation of wood.

and owing to its porous structure fresh charcoal can absorb many times its own volume of gas. Fresh charcoal also absorbs solids and liquids, and is therefore used for decolorizing crude brown sugar and similar substances that have to be purified and decolorized for use as food materials.

(i) Set up the apparatus shown in Fig. 19. Heat the sawdust in the hard-glass tube gently at first and then more strongly. Light the *wood-gas* as it escapes from the jet and notice the large volume of gas given off by a small quantity of wood. When there is no further change, allow the apparatus to cool and examine the distillate. Test the watery part with litmus paper. Examine the *wood-tar* and notice its characteristic smell (wood-tar is present in ordinary wood-smoke).

(ii) Half-fill a test-tube with the given coloured liquid (a 0·01 per cent. solution of eosin† or of methylene blue†). Add a few grams of your freshly-prepared *wood-charcoal* and shake vigorously for a few minutes. Filter and notice that the filtrate is practically colourless. After the filter has stopped dripping, add a few c.cm. of alcohol (methylated spirit) to the charcoal on the filter-paper and notice that a coloured solution comes through the filter.

LAMP-BLACK

Lamp-black is very finely divided *soot* obtained by burning substances rich in carbon (e.g. 'natural gas' from oil-wells) in a limited supply of air. It is used in paints, printers' ink, Indian ink, carbon paper,* typewriter ribbons, and gramophone records. Rubber, when mixed with lamp-black, becomes particularly suitable for making pneumatic tyres, and more than half the world's production of lamp-black is used in the motor-tyre industry.

ANIMAL CHARCOAL

Animal charcoal is made by distilling bones in iron retorts. Ammonia and bone-oil are given off, leaving a residue of carbon mixed with about 90 per cent. of calcium phosphate, the main

mineral constituent of bones. If this mineral matter is removed by dissolving in acid, very finely divided charcoal (bone black) remains behind. Animal charcoal is a very good decolorizing agent. Repeat Exp. (ii) in p. 57 with animal charcoal.

CRYSTALLINE FORMS OF CARBON

Graphite, which is used in making 'lead' pencils, is also known as 'black lead', because it was formerly supposed to contain lead. Natural graphite is found in Ceylon, Siberia, Poland, and North

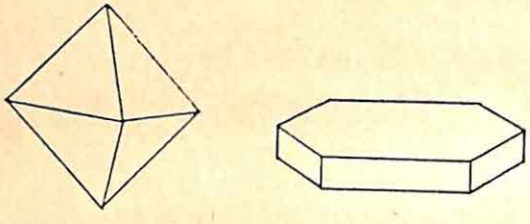


FIG. 20. Carbon crystals.

America. A very soft and pure artificial graphite is made at Niagara Falls, where coke or anthracite is heated very strongly in an electric furnace in the absence of air.

Graphite crystallizes in soft, flat plates that feel slippery, and it is therefore used for *lubricating* heavy machinery. The filling for lead pencils is made by mixing graphite and clay in suitable proportions and then heating the mixture. A 'soft' pencil has a larger proportion of graphite, while a 'hard' pencil has a larger proportion of clay.

Diamonds are found in many parts of the world, but chiefly in South Africa, South America, and India. For ornamental purposes, in jewellery, the natural diamonds are cut and polished so as to reflect as much light as possible. Diamond is the hardest substance known, and the splinters produced in cutting diamonds for jewellery are used for cutting glass. Black or other discoloured diamonds, which are useless for jewels, are mounted in the faces of rock-drills used in boring for oil and water. About 80 per cent.

of the world's annual diamond production is used for industrial purposes.

Artificial diamonds are said to have been made by dissolving carbon in liquid iron at a temperature of about 3,000° C. in an electric furnace, and then cooling suddenly. After cooling, the iron was dissolved away with acid, leaving very small diamonds behind, the largest being about 0·5 mm. in diameter. This process is of great scientific interest, but it has no commercial value.

COAL

Coal is an organic sedimentary rock formed from the highly compressed and partly decomposed remains of plants that once grew in swamps. We believe that the coal-beds that are now mostly buried in the Earth were once swamps on the surface. In these swamps grew forests of trees—the trunks and roots of some of them being still preserved as *fossils* in many coal-beds. As the plants died they fell into the water, where lack of oxygen prevented rapid decay. In time a deep *peat-bed*† was formed. Then, owing to some movement of the Earth's crust, the peat-bed became covered with a deep layer of weathered rock fragments* and in time both the peat and the overlying mud or silt were compressed and solidified into rock.

Brown coal (or *lignite*†) has not been compressed so strongly as ordinary black coal. It is less completely decomposed and contains less 'fixed' carbon and much more water than ordinary coal.

Soft coal (ordinary or *bituminous coal*†) is a shiny, black, brittle solid that burns with a luminous flame and much smoke. *Hard coal* (steam coal or *anthracite*†) is a dense, rock-like solid that burns slowly without flame or smoke, leaving very little ash. The different kinds of coal contain different proportions of carbon, e.g. brown coal contains 60–75 per cent. of carbon, soft coal has 75–90 per cent., hard coal has 90–93 per cent.

COAL-GAS

Set up the apparatus shown in Fig. 21, using a 3 in. $\times \frac{1}{2}$ in. hard-glass test-tube, half-full of powdered coal. Heat strongly and first

ignite the crude gas before connecting up the condenser and purifier. Notice the smoky flame. Then connect the apparatus to the condenser and purifier and set fire to the purified gas. Stop

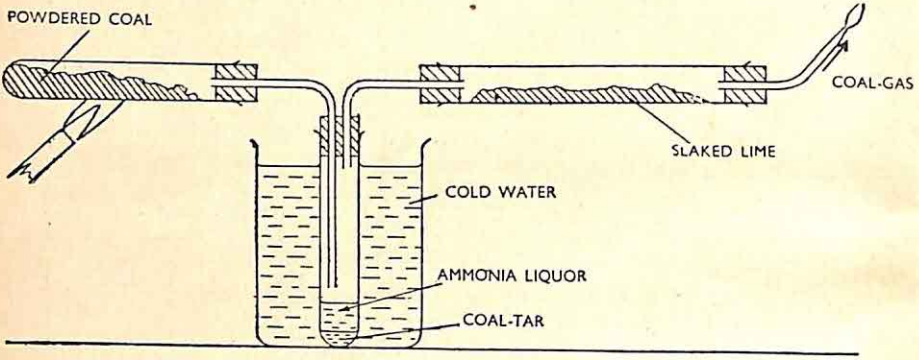


FIG. 21. Distillation of coal.

heating when no more gas is given off and allow the apparatus to cool. When cool, test the watery distillate with litmus paper. Shake out the residue from the hard-glass tube and examine this *coke*.

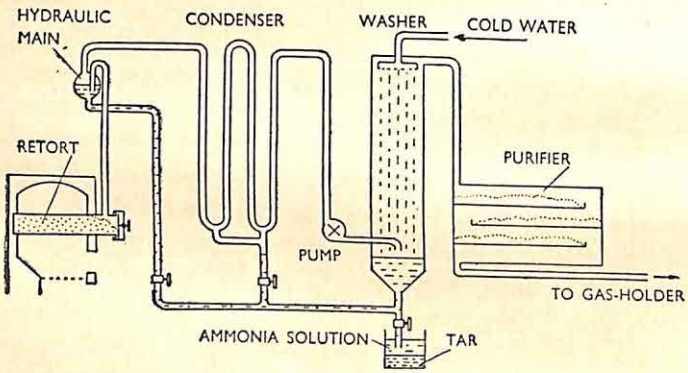


FIG. 22. Coal-gas works (diagrammatic).

Coal-gas is used very widely in many countries for heating and lighting. It is manufactured in a *gas-works* (see Fig. 22) by the *destructive distillation*† *of coal* in fire-clay retorts (i.e. by strongly heating coal in the absence of air). The coal-gas, mixed with many

impurities, passes through pipes into the *hydraulic main*,* a channel containing water that cools the gas and removes some of the impurities besides serving as a valve to stop gas escaping when a retort is opened. *Coke*, a porous form of carbon, is left in the retorts, and the crude* gas, containing *coal-tar*, *ammonia*, and *sulphur compounds*, passes through several stages of purification. It first passes through the *condensers*, where the *tar* and water, containing dissolved *ammonia*, are removed from the gas. The coal-gas next passes through a *washer*, where the remainder of the ammonia dissolves in water. Sulphur compounds are then removed by passing the gas through the *purifier*, over layers of iron oxide or slaked lime, which combine with the sulphur compounds. The purified coal-gas then passes into the *gas-holder*, ready for delivery through pipes to the users.

Coal-gas is a mixture of hydrogen (40–50 per cent.) and nitrogen (8 per cent.) with several gaseous compounds of carbon, the most important of these being methane (30–33 per cent.) and carbon monoxide (5–10 per cent.). It is this latter substance that makes coal-gas so poisonous. This mixture of gases burns in air with a smoky flame, but the flame is made hotter and less smoky (i.e. the combustion to carbon dioxide and water becomes complete) if air is mixed with the gas, as in a Bunsen burner. If too much air is introduced, an explosive mixture is produced.

BY-PRODUCTS FROM COAL-GAS MANUFACTURE

(i) *Coke*. The coke left behind in the retorts contains about 80 per cent. of carbon and is used as a cheap fuel, for making '*producer-gas*'† and '*water-gas*'† and also for reducing metals from their ores, particularly in the manufacture of iron in the blast furnace.

(ii) *Coal-tar*. Coal-tar is a thick black oily liquid that is a complicated mixture of at least 200 different substances, some of which are of great commercial value. The more valuable constituents are separated from the tar by *distillation*, condensing and collecting the vapour that comes off at different temperatures. The

first part of the tar to distil over is known as the 'light oil', containing *benzene*,† *toluene*,† and *xylene*,† which form the raw materials for the preparation of *synthetic dyestuffs* and *fine organic chemicals*. The second fraction to distil over is known as the 'middle oil', and this consists mainly of *phenol* (or *carbolic acid*†), and *naphthalene*. Carbolic acid is very poisonous to bacteria and is used as a disinfectant, while one use of naphthalene is as 'moth-balls' to drive away insects. The next fraction is the 'heavy oil', used for preserving wood. The last fraction to distil over is the 'anthracene oil', which yields *anthracene*,† another raw material for synthetic dyes, and the residue left behind in the retort is *pitch*.†

(iii) *Ammonia*. The ammonia in the liquid from the washers is set free and then absorbed in sulphuric acid, forming ammonium sulphate, which is used as a fertilizer for supplying crops with combined nitrogen. A ton of soft coal yields *approximately* 11,000 cu. ft. of coal-gas, 100 lb. of coal-tar, 20–40 lb. of ammonium sulphate, and 1,500 lb. of coke.

SOME COMPOUNDS OF CARBON

CARBON MONOXIDE

In Book Two, Chap. I, we spent some time studying carbon dioxide (CO_2). This, however, is not the only oxide of carbon; there is another important compound of carbon and oxygen—*carbon monoxide* (CO). This compound, which is also a gas, is like carbon dioxide in having neither colour nor smell, but it differs in being very inflammable and very poisonous.

Carbon monoxide is formed when carbon burns in a limited supply of air. It can be prepared by passing carbon dioxide through a tube containing red-hot carbon:

$$C + CO_2 \longrightarrow 2CO$$

After passing over the red-hot carbon, the gas is passed through sodium hydroxide solution, to remove any traces of carbon dioxide that have not been reduced by the carbon. The gas can be collected over water.

The same reaction takes place in a charcoal or coke fire (see Fig. 23), where carbon dioxide, formed at the bottom of the fire, is reduced to carbon monoxide in the middle of the fire, and this carbon monoxide combines with the excess of oxygen above the fire to form carbon dioxide.

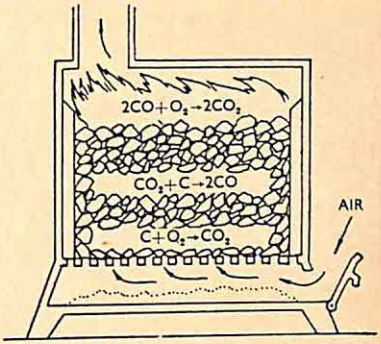


FIG. 23. Reactions in coke or charcoal fire.

PROPERTIES OF CARBON MONOXIDE

Carbon monoxide is a gas without colour, taste, or smell. It is slightly 'lighter' than air and is almost insoluble in water. Carbon monoxide burns readily in air with a characteristic blue flame, forming carbon dioxide:

$$2CO+O_2 \longrightarrow 2CO_2$$

(This blue flame is often seen above a charcoal fire.) Because of the readiness with which it combines with oxygen, carbon monoxide is often used as a *reducing agent*, e.g. in the blast furnace:

$$Fe_2O_3+3CO \longrightarrow 2Fe+3CO_2$$

Carbon monoxide is a very poisonous gas. As little as one volume of carbon monoxide mixed with 100,000 volumes of air will make people seriously ill, while 1 per cent. of carbon monoxide in air causes death in a few minutes, because the haemoglobin in the red blood-cells takes up carbon monoxide in preference to oxygen. Since carbon monoxide is formed by motor-engines running on a rich mixture (e.g. when starting up), it is very dangerous to run an engine inside a closed building and then breathe the exhaust gases.

PRODUCER-GAS AND WATER-GAS

When charcoal burns in a plentiful supply of air, it is completely oxidized to carbon dioxide (see Equation 1, p. 64). If, however, the air is supplied to a bed of red-hot charcoal at one point only,

the carbon dioxide has to pass through some of the charcoal, and it is reduced to carbon monoxide (see Fig. 23 and Equation 2 below). Both reactions (1) and (2) liberate heat.

$$(1)\ C+O_2 \longrightarrow CO_2$$
$$(2)\ CO_2+C \longrightarrow 2CO$$

The nitrogen that enters with the air passes through unchanged. The mixture of carbon monoxide and nitrogen that comes from the furnace is called *producer-gas*. The carbon monoxide in it burns with a hot flame (see Equation 3). This inflammable producer-gas has been used as a substitute for petrol vapour in some motor-car engines.

$$(3)\ 2CO+O_2 \longrightarrow 2CO_2$$

As commercial charcoal normally contains some *water*, another reaction (Equation 4) also takes place, giving carbon monoxide and hydrogen.

$$(4)\ C+H_2O \longrightarrow CO+H_2$$

This mixture is called *water-gas*. It burns (see Equations 3 and 5) with the production of about three times as much heat as producer-gas, and, in some 'gas producers', water is allowed to drip on a hot plate near the entrance to the furnace so that a mixture of water-gas and producer-gas is formed.

$$(5)\ 2H_2+O_2 \longrightarrow 2H_2O$$

The advantages of this addition of water are twofold: Firstly, it saves fuel because hydrogen, from the water, is added to the mixture by reaction (4); secondly, it cools the fire because reaction (4) absorbs heat instead of liberating it as do reactions (1) and (2).

A 'gas-producer' can be fitted to a motor-vehicle and charcoal carried instead of petrol. The charcoal is ignited in a stove and a stream of air is drawn through the fire by the engine, which is usually started on petrol. The producer-gas then passes into a settling chamber in which the larger solid impurities are removed. It is then cooled and filtered through graded cloth screens and mixed with the right amount of air in a valve that replaces the

petrol carburettor. The mixture is then exploded in the engine-cylinders in the usual way. Producer-gas contains a good deal of inactive nitrogen, and it does not give as much power as petrol, even when enriched with water-gas.

Producer-gas and water-gas are also used for industrial purposes, but are usually made from coke and not from charcoal. Coal-gas is sometimes mixed with producer-gas and water-gas to produce an industrial fuel-gas of high calorific value. Much of the hydrogen used for industrial processes is extracted from water-gas.

HYDROCARBONS

There are many compounds containing only *carbon* and *hydrogen*, called *hydrocarbons*, and, as we might expect, they all burn readily, giving out light and heat.

METHANE

Methane (or *marsh-gas*†) is the simplest hydrocarbon and is formed in Nature by bacterial action on organic matter in the absence of air, e.g. under water. These bacteria produce enzymes that *hydrolyse*† (see p. 76) cellulose to methane and carbon dioxide:

$$(C_6H_{10}O_5)_n + nH_2O \longrightarrow 3nCH_4 + 3nCO_2$$

Bubbles of methane can sometimes be seen rising to the surface of marshy pools. This gas also escapes from cracks in coal-mines, where it is known to the miners as '*fire-damp*',† causing dangerous explosions, for a mixture of methane and air explodes violently when ignited:

$$CH_4 + 2O_2 \longrightarrow CO_2 + 2H_2O$$

Since one-fifth of the air is oxygen, one volume of methane uses up all the oxygen from 10 volumes of air, forming carbon dioxide in its place. Hence there is a serious danger of suffocation by carbon dioxide (called '*after-damp*'† by miners) after a 'fire-damp' explosion in a coal-mine has removed most of the oxygen.

Methane is also present in the *natural gas* that collects above petroleum deposits underground (see Fig. 25). It is also formed

during bacterial decomposition of sewage, in the 'sludge*-digestion' tanks at a sewage-disposal works (see Book IIIA, Chap. 8).

THE PARAFFIN SERIES OF HYDROCARBONS

There is a whole series of hydrocarbons similar to methane, e.g. *methane*—CH_4, *ethane*—C_2H_6, *propane*—C_3H_8, *butane*—C_4H_{10}, *pentane*—C_5H_{12}, *hexane*—C_6H_{14}, and so on. This series of hydrocarbons is called the *paraffin series*, and has the general formula C_nH_{2n+2}. The lowest members of the series (CH_4 to C_4H_{10}) are gases at ordinary temperatures, those from C_5H_{12} to $C_{17}H_{36}$ are liquids or semi-solids, while the paraffins from $C_{18}H_{38}$ upwards are solids; e.g. paraffin-wax is a mixture of solid paraffins with very large molecules.

Several of these paraffin hydrocarbons are present in *natural gas*,† e.g. an average sample of American oil-field gas contains about 35 per cent. of methane, 30 per cent. of ethane, and 22 per cent. of propane. These are all valuable fuels, and pipe-lines carry this natural gas from the oil-fields to towns many hundreds of miles away. Places out of reach of such pipe-lines can be supplied with '*bottled gas*', prepared from natural gas and compressed to a liquid in steel cylinders.

THE ETHYLENE SERIES OF HYDROCARBONS

Ethylene,† C_2H_4, is the simplest member of another series of hydrocarbons called the *ethylene series*. Ethylene is present in coal-gas, burning with a luminous, smoky flame. The general formula of this series is C_nH_{2n}. The hydrocarbons of the ethylene series all burn with a more luminous and more smoky flame than the corresponding paraffins, since they contain a greater proportion of carbon.

THE ACETYLENE SERIES OF HYDROCARBONS

Acetylene, C_2H_2, is the first member of another series of hydrocarbons called the *acetylene series*, the general formula of the series

being C_nH_{2n-2}. Acetylene is easily prepared by dropping water on calcium carbide, a method used in some lamps:

$$CaC_2 + 2H_2O \longrightarrow C_2H_2 + Ca(OH)_2$$

Acetylene is a colourless gas with an unpleasant smell. It burns in air with a luminous and very smoky flame. Special burners are made, however, that will burn acetylene with a bright white light and without smoke. When mixed with air or oxygen, acetylene explodes with great violence if a light is applied to the mixture:

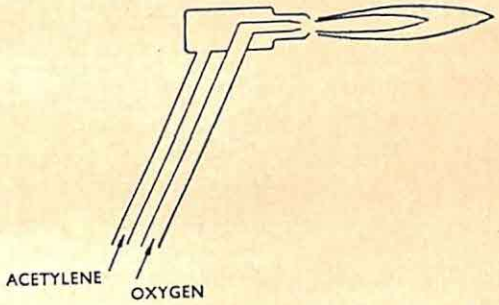


FIG. 24. Oxy-acetylene blowpipe (diagrammatic).

in fact, it is not safe for you to try the experiment. The *oxy-acetylene*† *blowpipe*, burning acetylene in oxygen, provides the hottest flame known (over 3,000° C.). It will 'cut' through steel quite easily. The metal is first oxidized and the oxide then melts and falls away as 'scale'. The oxy-acetylene flame is also used for *welding* iron and steel (see Fig. 24).

THE BENZENE SERIES OF HYDROCARBONS

Benzene—C_6H_6—is the first member of another important series of hydrocarbons, called the *benzene series*. Higher members of this series are *toluene* and *xylene*, which are present in coal-tar along with benzene. (N.B.—Do not confuse benzene with benzine.*)

You will realize from this outline that the number of different compounds of carbon and hydrogen is very large indeed, but the study of all these hydrocarbons is simplified by the fact that there

are not very many series and that *members of the same series all have very similar properties.*

PETROLEUM

Petroleum (*mineral-oil*, or *rock-oil*) is a natural mixture of various hydrocarbons. It is found in many parts of the world, e.g. North America, Russia, Rumania, Persia, Iraq, Trinidad, and South-East Asia. American petroleum consists mainly of *paraffins*, i.e. hydrocarbons of the *methane series*, while Iraqi petroleum contains hydrocarbons of the *benzene series.*

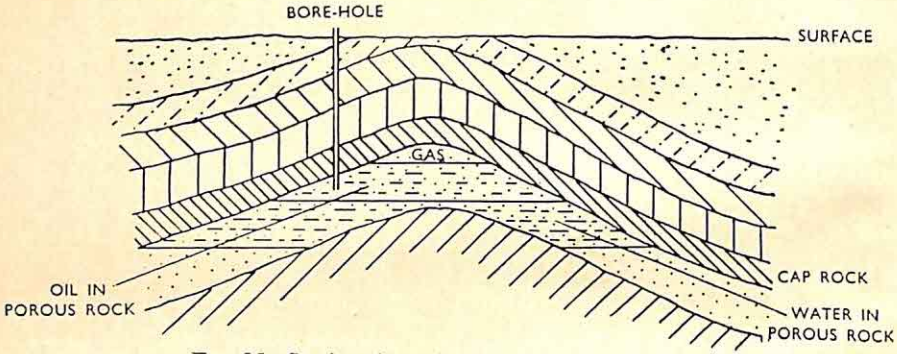


FIG. 25. Section through oil-well (diagrammatic).

Petroleum is obtained by drilling a bore-hole down through the earth until the petroleum is reached (see Fig. 25). At first, the oil usually rushes out under its own pressure, but the pressure decreases slowly until, after a time, the oil has to be pumped to the surface. The crude petroleum is a thick, heavy liquid that varies in colour from greenish-brown to black, consisting of a mixture of several hundred different hydrocarbons. The purification, or *refining*, of this crude petroleum and its products is now a great industry, yielding several hundred commercial products.

PETROLEUM REFINING

The constituents of petroleum are separated by *fractional distillation*, each constituent having a different boiling-point and thus distilling over at a definite temperature. The crude petroleum is

run into large *stills** and then heated gradually. The fraction with the lowest boiling-point, which distils over first, is the main source of *petrol* (or gasoline†). The next product to distil over is *kerosene* (or paraffin-oil), which has a higher boiling-point and catches fire less readily than petrol. At a still higher temperature, large quantities of a thicker and heavier oil distil over, known as *gas-oil*,† which is used as a fuel for oil-engines (e.g. Diesel-engines), and which can also be converted into petrol by a process known as '*cracking*'.† (In this process the gas-oil is passed through a red-hot tube at a high pressure, and the 'heavy' hydrocarbons of high molecular weight and high boiling-point split up into 'light' hydrocarbons of lower molecular weight and lower boiling-point. About half the world's supply of petrol is now made by 'cracking' heavier oils.) At a still higher temperature, a very thick oil distils over that yields *lubricating oils*, and also *vaseline* (petroleum-jelly) and *paraffin-wax*. *Pitch* is left behind in the still, together with some *coke*. A typical American petroleum yields about 4 per cent. of gas, 25 per cent. petrol, 15 per cent. kerosene, 40 per cent. gas-oil, 12 per cent. lubricating oil, and 4 per cent. coke.

COMBUSTION AND FLAME

A chemical change that is accompanied by the production of heat and light is called combustion, as when we speak of the combustion of coal in a railway-engine; the combustion of petrol inside the cylinder of a motor-engine (internal combustion); the combustion of magnesium in oxygen; the combustion of copper in chlorine. Some combustion takes place without *flame*; in fact, flame is only an accidental result of combustion. In the combustion of iron in sulphur, for example, there is no flame; the mixture merely becomes white-hot. *Substances that yield gases on heating, e.g. burning wood or oil, burn with a flame.*

We can distinguish between two types of flames, (i) *luminous flames*, which give out light, e.g. the candle flame, and (ii) *non-luminous flames*, which produce little or no light, e.g. burning hydrogen or alcohol, or a well-regulated Bunsen flame.

The wax of which the candle is made cannot burn until it is changed into gas, and this is the function of the *wick*.* The heat of the flame changes some solid wax into liquid wax, and this liquid wax rises up the wick by *capillarity*, until it reaches the inside of the flame, where it changes into gas.

The heat of the flame then decomposes the hydrocarbons of which the wax is composed, setting free tiny particles of carbon that are heated white-hot by the hydrogen as it burns to form

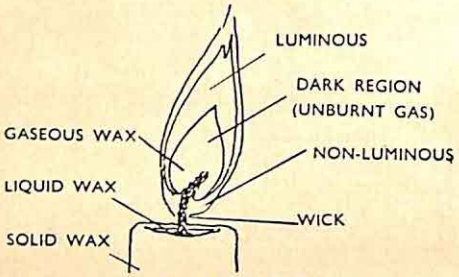


FIG. 26. Candle flame.

steam. If there is a good supply of air, the carbon near the outside of the flame burns in the oxygen of the air, forming carbon dioxide, but if the air supply is insufficient, some of the carbon escapes, unburnt, as black smoke (or soot).

When viewed in a darkened room, a candle flame (see Fig. 26) is seen to consist of (i) a dark, inner region of unburnt gas, (ii) a luminous region containing white-hot carbon particles, and (iii) a non-luminous outer region (which can be seen more clearly by doing a sodium flame-test).

Hold a sheet of paper horizontally across a candle flame, withdrawing it just before it catches fire. Notice the charred ring, showing that the middle of the flame is cool, consisting of unburnt gas. Make a 'vertical section' of the flame in the same way.

Hold a piece of broken porcelain in the candle flame. It is blackened with *soot*, formed by cooling the white-hot carbon particles in the luminous region of the flame.

Blow out a lighted candle and then *quickly* bring a flame to a point about half an inch above the wick. The gas rising from the hot wick ignites.

THE BUNSEN FLAME

For laboratory purposes we require a hot flame that will not deposit soot, and the Bunsen burner provides such a flame. When the air-holes are closed, we get a luminous, smoky flame. When the air-holes are opened, the rush of gas through the jet at the bottom of the burner-tube draws in air through the air-holes, and this air mixes with the gas in the burner-tube. With the right amount of air, all the carbon in the gas burns at once to carbon dioxide, instead of separating as white-hot particles as it does when the air-holes are closed. Hence, when the air-regulator admits the correct amount of air, the coal-gas burns with a clean greenish-blue flame called the *Bunsen flame*. If the proportion of air is increased, an explosive mixture is formed and the flame 'strikes back' down the burner-tube, burning at the jet itself.[1] This 'striking back' can be prevented by covering the top of the burner-tube with wire gauze. The principle involved is the same as in the miners' safety lamp, which we shall discuss shortly.

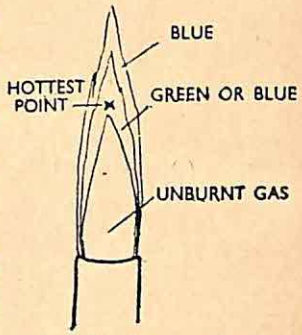


FIG. 27. Bunsen flame.

The Bunsen flame consists of two regions of burning gas surrounding an inner cone of the unburnt gas (see Fig. 27). If a piece of paper is held for a moment horizontally across the flame, the paper is charred in a ring, showing that the inside of the flame is quite cool, unburnt gas. In the same way, if a wide tube is connected to the gas supply and the top is covered with wire gauze, a small heap of gunpowder on the centre of the gauze does not catch fire when the gas is lighted. This shows clearly that the flame is *hollow*.

[1] With petrol-gas, however, the flame usually 'floats off' when too much air is admitted.

The inner region of the Bunsen flame is green or blue in colour, and in this part of the flame the hydrocarbons in the gas are partly burnt to form carbon monoxide and steam. The outer region is blue in colour, and in this part of the flame the carbon monoxide formed in the inner cone burns to form carbon dioxide.

The Bunsen flame gives out much heat but little light. For lighting purposes, this heat energy can be converted into light energy by allowing a Bunsen flame to heat a 'gas-*mantle*',† which

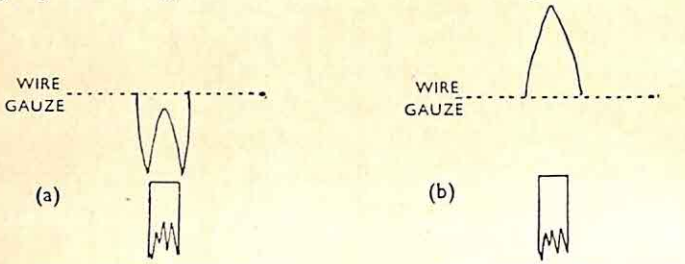


FIG. 28. Flame sieves.

contains the oxides of *thorium*† (99 per cent.) and of *cerium*† (1 per cent.). These oxides when heated become white-hot, and the mantle gives out a bright, white light.

THE SAFETY LAMP

When candles were used in coal-mines, 'fire-damp' (methane) explosions were very common and many lives were lost. In 1815, Sir Humphry Davy studied the causes and possible means of preventing these accidents. He found that if a flame is cooled, it is extinguished, because inflammable gases all have a definite temperature below which they cannot be ignited. This temperature is called the '*ignition temperature*'. The ignition temperature of methane is fairly high, so that it cannot be set alight by a red-hot object, but only by something at a higher temperature. Davy also noticed that if a piece of wire gauze is held over a flame, the flame burns on the underside of the gauze only (see Fig. 28(*a*)). The flame does not pass through the gauze because the metal conducts away

the heat and does not get hot enough to ignite the gas passing through the gauze. If, however, the gauze is left over the flame for a long time, it may become hot enough to ignite the gas above the gauze. Conversely, if a piece of wire gauze is held over an

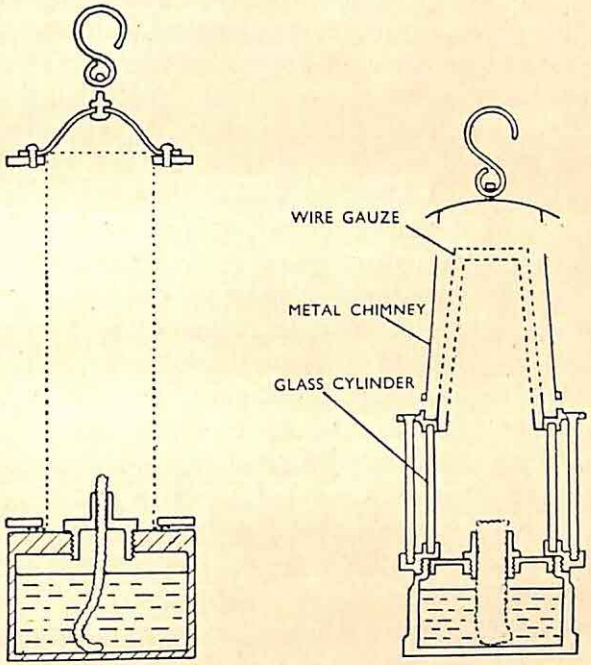


FIG. 29.

Original form. Modern form.

unlighted burner, the gas can be lighted *above* the gauze, but the flame does not pass down through the gauze and ignite the gas at the top of the burner-tube (see Fig. 28(*b*)).

Experiments like these led Davy to the idea of using a wire gauze as a '*flame sieve*', which, while allowing gas, air, and light to pass, would stop the passage of a flame. The original '*Davy safety lamp*' consisted of an oil-lamp surrounded by a cylinder of wire gauze (see Fig. 29). When placed in air containing 'fire-damp',

the methane passes through the gauze and burns *on the inside* of the gauze without igniting the explosive mixture outside.

Later and improved forms of the Davy lamp have a strong glass cylinder round the lower part of the lamp, and gauze above (see Fig. 29). This allows more light to escape and also prevents the flame from being blown through the gauze by a wind. (In modern coal-mines, a rapid current of fresh air is blown through the workings to dilute the methane with so much air that the mixture is no longer explosive.)

CARBOHYDRATES

Carbohydrates are neutral compounds of carbon, hydrogen, and oxygen, with the hydrogen and oxygen present in the same relative proportions as in water, i.e. 2 atoms of hydrogen to every atom of oxygen. The carbohydrates, which play a very important part in plant and animal life, can be classified as (i) *sugars*, (ii) *starches*, (iii) *celluloses*; or for chemical purposes as (*a*) *simple sugars* (or monosaccharides,† e.g. glucose and fructose†), (*b*) *complex sugars* (or disaccharides,† e.g. sucrose,† lactose,† and maltose†), and (*c*) *complex carbohydrates* (e.g. starch, dextrin,† glycogen, and cellulose).

SIMPLE SUGARS—MONOSACCHARIDES

Glucose ($C_6H_{12}O_6$—grape sugar or dextrose) is the most important of the simpler sugars. It is probably the first carbohydrate formed during photosynthesis in green plants and it is always present in plant cells, where it may be used in building proteins and also in respiration. It is found in quantity in some fruits, e.g. grapes (raisins—dried grapes—contain hard crystals of glucose). The blood always contains small quantities of glucose which is carried in the blood to the body cells, where it is oxidized, yielding *energy*. Commercially, glucose is made by boiling starch with very dilute acids under pressure. The product provides a cheap sweetening agent used in making jams and sweets.

A closely related simple sugar is *fructose*, common in fruits and

in nectar from flowers. The simple sugars are sweet, crystalline substances that are very soluble in water. They cannot be split up into simpler carbohydrates, but they are readily oxidized and therefore act as *reducing agents*. The usual test for glucose is based upon this fact:

Fehling's Test.—Take 2–3 c.cm. of 1 per cent. glucose solution (1 in. in a $\frac{5}{8}$-in. test-tube) and add an equal volume of diluted (1/5) Fehling's solution (an alkaline solution of copper sulphate containing also Rochelle salt†—sodium hydrogen tartrate†). Place the test-tube in a boiling water-bath. A red precipitate of cuprous oxide shows the presence of a reducing sugar.

COMPLEX SUGARS—DISACCHARIDES

Ordinary household 'sugar' is *sucrose* (or *cane-sugar*) and it is obtained from the sugar-cane,† the sugar-beet,† the sweet sorghum-cane,† or the sugar-palm.† Sugar-cane is crushed between iron rollers to extract the cane-juice. This juice, which contains about 20 per cent. of sugar, is then purified, filtered, and evaporated, and the crystals of 'raw sugar' are removed from the liquid residue (treacle† or molasses†). This crude brown sugar is refined by washing the brown crystals and then dissolving them in hot water. This brown liquid is filtered and then decolorized with charcoal. It is then evaporated under reduced pressure at a low temperature, to produce the familiar white crystals used in the household. Sugar-beets are crushed to a pulp and the sugar is extracted with water, the solution being purified and evaporated by processes similar to those used in the manufacture of sugar from the sugar-cane. Sugar is obtained from the sugar-palm by tapping the flower-shoots and evaporating the sap that flows out. The product, in each case, is *sucrose*.

Sucrose, $C_{12}H_{22}O_{11}$, is a complex sugar (a disaccharide) because it can be split up into two simple sugars by boiling with very dilute acids or by the action of certain *enzymes* (see p. 107), one molecule of sucrose yielding one molecule of glucose and one of fructose:

$$C_{12}H_{22}O_{11} + H_2O \longrightarrow C_6H_{12}O_6\ (\textit{glucose}) + C_6H_{12}O_6\ (\textit{fructose})$$

(This is an example of *hydrolysis*, the chemical decomposition of a substance by water, the water itself being also decomposed.)

The mixture of glucose and fructose, produced by boiling sucrose with dilute acid, is called '*invert sugar*',† and it is more readily digested than sucrose. *Honey* is a natural form of invert sugar. *Jam* (particularly 'home-made' jam) contains a large proportion of invert sugar formed when the original sucrose is boiled with the acid fruit-juice.

Another common complex sugar is *lactose* (or *milk-sugar*), which is present in cows' milk to the extent of about 5 per cent. Like sucrose, lactose can be hydrolysed into two simple sugars, one molecule of lactose yielding one molecule of glucose and one molecule of galactose.† Lactose is not sweet like sucrose, but is almost tasteless.

Maltose (or *malt-sugar*), another complex sugar, is formed in germinating seeds by the action of the enzyme *diastase*† on the starch stored as a reserve food material. (The enzyme *ptyalin*† in the saliva also hydrolyses starch into maltose.) Commercially, *barley* is made into *malt*† by allowing the grains to germinate until some diastase has been formed. Germination is then stopped by heating, and the malted barley is extracted with water, yielding 'malt extract', which contains maltose and diastase. Maltose, like sucrose and lactose, is readily hydrolysed by dilute acids or by a suitable enzyme (e.g. *maltase*,† which is present in the human small intestine), yielding two molecules of glucose.

EXPERIMENTS WITH COMPLEX SUGARS—DISACCHARIDES

There are two kinds of disaccharides, (*a*) those that reduce copper solutions, e.g. maltose and lactose, and (*b*) those that do not reduce, e.g. sucrose. The reducing power of glucose and fructose, however, is greater than that of maltose and lactose.

Repeat Fehling's test with maltose, lactose, and sucrose, and compare their reducing power with that of glucose and fructose. Take five test-tubes, A, B, C, D, and E, each containing 2–3 c.cm. of diluted Fehling's solution (1/5). To A add 1 c.cm. of 1 per cent.

glucose solution, to B add 1 c.cm. of 1 per cent. fructose solution, to C add 1 c.cm. of 1 per cent. maltose solution, to D add 1 c.cm. of 1 per cent. lactose, and to E add 1 c.cm. of 1 per cent. sucrose. Place all five test-tubes in a boiling water-bath and notice when a precipitate of cuprous oxide appears: first in A (glucose) and B (fructose) and a little later in C (maltose) and D (lactose). Notice that *no precipitate forms in* E (sucrose). *Sucrose has no reducing action*, and the reducing power of glucose and fructose is greater than that of maltose and lactose.

To convert Complex Sugars into Simple Sugars.—Take 2–3 c.cm. of (*a*) 1 per cent. sucrose solution, (*b*) 1 per cent. maltose solution, and (*c*) 1 per cent. lactose solution, and add to each tube 1 c.cm. of dilute hydrochloric acid (2N). Boil for 2 minutes. Cool, and then neutralize with 1 c.cm. of dilute sodium hydroxide solution (2N). Take four test-tubes, A, B, C, and D, and put 2–3 c.cm. of diluted Fehling's solution in each. To A add 2 c.cm. of the boiled sucrose solution; to B add 2 c.cm. of the boiled maltose solution; to C add 2 c.cm. of the boiled lactose solution; and to D add 2 c.cm. of 1 per cent. glucose solution, unboiled. Place all four tubes in a boiling water-bath and notice the rapid reduction to red cuprous oxide in every case, showing that the complex sugars sucrose, lactose, and maltose have been hydrolysed into simple sugars with greater reducing power.[1]

COMPLEX CARBOHYDRATES—POLYSACCHARIDES†

Starch $(C_6H_{10}O_5)_n$ is one of the commonest and most important of the complex carbohydrates. It is the chief reserve food material of plants, in which carbohydrates, made in the leaf as *glucose*, are transported as soluble *sucrose* and stored in the form of insoluble *starch*.

Starch is always formed in plants as solid grains in special dense parts of the protoplasm called the *plastids*. In the leaf these are

[1] Only a selection of experiments is given in this book to illustrate chemical physiology. Further experiments are given in the Teacher's Handbook, together with full details of the preparation and composition of test solutions and reagents.

green, and starch formed in them by day, as tiny grains, disappears again during the night, being changed to sugar. In stems, roots, and storage organs, e.g. English potatoes, starch forms larger and more lasting grains in colourless plastids.

Starch grains are *colloidal*† solids (see p. 118) quite insoluble in water until they have been heated or subjected to enzyme action. Then they break down into smaller particles that go into suspension in water, forming a translucent paste (but not a true solution). If this starch-paste is heated with acid or allowed to stand with the enzymes *diastase* or *ptyalin*, the starch is hydrolysed to glucose:

$$(C_6H_{10}O_5)_n + nH_2O \longrightarrow nC_6H_{12}O_6$$

Dextrin is another complex carbohydrate formed when starch is baked at about 200° C. Dextrin is also formed as an intermediate product when starch is heated with acids or treated with enzymes, e.g. ptyalin. It is a yellow or brown solid that dissolves in water. Dextrin is used commercially for making gum,† e.g. that used on postage stamps.

Glycogen (sometimes called '*animal starch*') is the commonest carbohydrate reserve in animals, where it is found mainly in the liver and the muscles. (A well-fed man has about 200 gm. of glycogen in his liver and about 350 gm. in his muscles.)

Glycogen is difficult to extract from liver and muscle since it is rapidly changed into sugar after the animal dies. It is most readily obtained for laboratory experiments from oysters and other molluscs. Glycogen is soluble in water, forming a translucent solution.

Cellulose is the name given to a number of complex carbohydrates that form the cell-walls of plants (see also p. 98). Cotton, linen, and filter-paper are nearly pure cellulose. Cellulose, in its various forms, is of very great importance to man. It is one of the main constituents of wood, which is used in enormous quantities for a great variety of purposes. Paper, too, is mainly cellulose. The cheapest quality of paper ('newsprint') is made from wood-pulp, which is heated with chemicals, leaving cellulose. Better

quality paper is made by mixing the wood-pulp with china-clay or some similar 'filler' and then rolling the mixture into sheets. The best quality paper is made from old linen, the cellulose fibres in linen being longer than those in wood-pulp, thus producing a tougher paper. Paper passed quickly through fairly concentrated sulphuric acid becomes water-proof 'parchment* paper'.

EXPERIMENTS WITH COMPLEX CARBOHYDRATES—POLYSACCHARIDES

Starch. (i) Take about 1 c.cm. of 1 per cent. starch solution in a test-tube and add one drop of iodine solution (N/20 in potassium iodide). Notice the deep blue-black colour. Fill up the tube with water and notice that the colour is still deep blue.

(ii) Take about 1 c.cm. of 1 per cent. starch solution and one drop of iodine solution, add about 3 c.cm. of water and boil. Notice that the blue colour disappears on heating. Cool under the tap. Notice that the blue colour returns on cooling. (N.B.—Remember that the iodine test does not work in hot solutions.)

(iii) Take 1 c.cm. of 1 per cent. starch solution and make it alkaline to litmus with sodium carbonate solution. Add a drop of iodine solution. There is no blue colour. Repeat this with 1 c.cm. of starch solution that has been acidified with dilute hydrochloric acid: iodine gives the usual blue colour. (N.B.—Remember that the iodine test does not work in alkaline solutions, but only in neutral or acid solutions.)

(iv) Take a small quantity of dry rice-starch in a test-tube and shake vigorously with about 5 c.cm. of cold water. Filter half the mixture. Add a drop of iodine solution to the filtrate and notice that no blue colour develops, showing that starch is insoluble in cold water. Boil the remainder of the suspension and then filter. Add one drop of iodine solution to the filtrate (after cooling). Notice that a blue colour is produced at once, showing that starch is soluble in hot water.

Dextrin. Take 2–3 c.cm. of a 1 per cent. solution of dextrin and add one drop of iodine solution. Notice the reddish-brown colour. (Compare this with the colour produced by adding one

drop of iodine to 2–3 c.cm. of water.) Heat the mixture of dextrin and iodine—the reddish-brown colour disappears. Cool under the tap and notice that the colour returns.

Glycogen. Take two test-tubes, A and B. In A put about 5 c.cm. of the given glycogen solution, and in B put about 5 c.cm. of water. Add equal quantities of iodine solution to each tube drop by drop and compare the colours. The reddish-brown colour in A is due to the presence of glycogen. (N.B.—The iodine test gives similar results with dextrin and glycogen. *Dextrin* solutions, however, are clear and *transparent*, while *glycogen* solutions are *translucent*.)

Cellulose. Put one drop of iodine on a piece of filter-paper. No blue colour is formed. Add one drop of sulphuric acid (50 per cent.) to the same spot; it produces a blue stain.

TO CONVERT STARCH INTO SIMPLER CARBOHYDRATES

(i) First show that starch has no reducing action. Take 2–3 c.cm. of 1 per cent. starch solution and add an equal volume of diluted Fehling's solution. Place in a boiling water-bath. No precipitate of cuprous oxide is formed, showing that *starch has no reducing action* on Fehling's solution.

(ii) Take about 10 c.cm. of 1 per cent. starch solution in a test-tube and add 1 c.cm. of concentrated hydrochloric acid. Place in a boiling water-bath, noting the time. On a white tile* (or a sheet of glass over white paper) put 12 drops of iodine solution and, at the end of every minute, quickly stir the starch solution with a glass rod and then remove a drop of the solution and test it with a drop of iodine on the white tile (washing the glass rod under the tap between successive tests). Notice the gradual change of colour from blue (starch) to purple and then reddish-brown (dextrin); finally the solution becomes colourless. Pour 2–3 c.cm. of the remaining solution into a test-tube and neutralize with about 1 c.cm. of sodium hydroxide solution (2N). Then add an equal volume of diluted Fehling's solution and heat in a boiling water-bath. A red precipitate of cuprous oxide shows that a reducing sugar (glucose in this case) has been formed.

(iii) It is interesting to compare the hydrolysis of starch by acids with that brought about by the enzyme *ptyalin*. Put 2–3 c.cm. of 1 per cent. starch solution in a test-tube and add a little saliva. Shake up and stand it in the 37° C. water-bath for a few minutes. Then add an equal volume of diluted Fehling's solution and heat in a boiling water-bath. A red precipitate of cuprous oxide shows the presence of a reducing sugar.

TO CONVERT GLYCOGEN INTO SIMPLER CARBOHYDRATES

(iv) Take about 5 c.cm. of the given glycogen solution and add a few drops of concentrated hydrochloric acid. Heat in a boiling water-bath for 10 minutes, then cool under the tap and divide into two parts A and B. To A add two or three drops of iodine solution. No reddish-brown colour develops, showing that the glycogen has disappeared. Neutralize B with sodium hydroxide solution, add an equal volume of diluted Fehling's solution, and heat in a boiling water-bath. A red (or yellow) precipitate shows that the glycogen has been converted into a reducing sugar (glucose in this case).

(v) If time permits, try the action of saliva on glycogen (see experiment (iii) above).

ALCOHOLIC FERMENTATION†

It has been known for centuries that when grape-juice or any other dilute solution of sugar is left exposed to the air, it loses its sweetness, acquires a burning taste, and becomes intoxicating.* The sugar is said to have *fermented* and it has been converted into *ethyl alcohol*,† giving off carbon dioxide in the process:

$$C_6H_{12}O_6 \longrightarrow 2C_2H_5OH + 2CO_2$$

The fermentation is brought about by *yeast*, a microscopic, non-green plant that consists of a single cell (see Fig. 30). Instead of seeds, the yeast-plant (like other *fungi*) forms *spores*, which are small enough and light enough to float about in the air, so that whenever sugar solutions are exposed to the air, yeast spores soon

enter, grow, and multiply, setting up fermentation. If this 'living dust' is kept out, sugar solutions cannot ferment. The actual chemical change involved in fermentation is brought about by an *enzyme* (called *zymase*†) contained in the yeast-cells, which acts as an organic catalyst. Living yeast uses the energy set free in

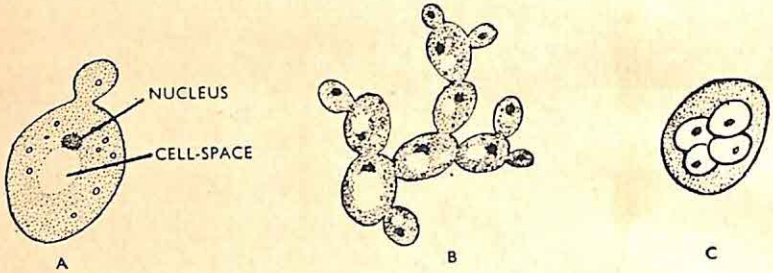


FIG. 30. Yeast. A. A single yeast plant with 'bud' (×2,500). B. Yeast plants reproducing by 'budding'. C. Four yeast spores inside a resting cell.

the process of *fermentation*, just as we use the energy set free by the more usual type of *respiration*.

Fermentation stops when about 14 per cent. of alcohol has been formed in the solution, and stronger alcohol is then obtained from this dilute alcohol by *fractional distillation*; the alcohol, having a lower boiling-point than water, distils off first. The strongest and purest alcohol is called *absolute alcohol* and contains over 99 per cent. of ethyl alcohol. In order to discourage people from drinking cheap alcohol sold for industrial purposes, crude wood-alcohol (methyl alcohol), paraffin, bone-oil, and dyes are added, the mixture being known as *methylated spirit*. Ethyl alcohol is used in large quantities as a solvent for making varnish and it can also be used, mixed with petrol, as a motor fuel.

Just as there is a *paraffin series* of hydrocarbons (methane—CH_4, ethane—C_2H_6, etc.), there is a corresponding series of *simple alcohols*, e.g. *methyl alcohol*—CH_3OH, *ethyl alcohol*—C_2H_5OH, and so on. In everyday use, when we speak of 'alcohol' we mean 'ethyl alcohol', as this is the commonest alcohol. 'Wood-alcohol' is 'methyl alcohol' and this is a very poisonous substance.

TO SHOW THE ALCOHOLIC FERMENTATION OF GLUCOSE

(*Demonstration.*) In a 500 c.cm. boiling-flask put 250 c.cm. of a 10 per cent. solution of glucose. Add 5 gm. of dried yeast and fit the flask with a cork carrying a glass tube leading to a tube containing clear lime-water, as shown in Fig. 31. This tube is closed by a two-holed cork carrying a soda-lime tube to absorb any carbon dioxide entering from the air. Keep the flask at about 37° C. for 24 hours (or, if this is not possible, first warm the flask to this temperature and afterwards keep in a warm room for at least 24 hours). Notice that large volumes of gas escape through the lime-water, turning it chalky at first, but it afterwards becomes clear again owing to the large volume of carbon dioxide that is given off (see Book Two, Chap. I).[1] After a day or two, filter the contents of the flask into a distilling-flask and distil over a small flame, collecting the first 25 c.cm. to come over. Take about 5 c.cm. of this distillate in a test-tube and add a few drops of sodium hydroxide solution (2N). Then add iodine solution, drop by drop, until the mixture is faintly brown. Warm gently, and notice the peculiar smell. Repeat this test with a few drops of ethyl alcohol. The same smell is produced. This is due to the compound *iodoform*,† produced from alcohol by the action of the alkali and the iodine. Repeat the test with a few c.cm. of 10 per cent. glucose solution—there is no smell of iodoform. Hence the yeast has converted some of the sugar into alcohol, setting free carbon dioxide during the process.

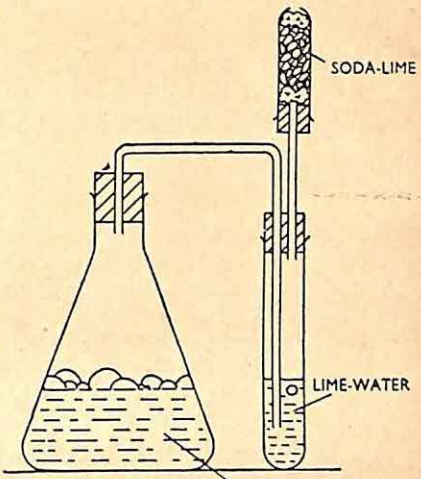


FIG. 31. Fermentation of glucose.

[1] Baryta-water—barium hydroxide† solution—may be used instead of lime-water as it does not become clear again when carbon dioxide is passed for a long time.

If you examine, under the high power of the microscope, a drop of the liquid from the flask, where the yeast has had a good supply of food for over 24 hours, you may see the yeast '*budding*', i.e. reproducing by division (see Fig. 30).

THE USES OF YEAST

Yeast has been used for centuries in the production of wines, beer, and spirits, and in bread-making.

Wine is made by fermenting the sweet juice of grapes with the help of various wild yeasts that live on the fruits. The fermentation produces ethyl alcohol, and the various kinds of wine differ according to the methods adopted, the type of grape, soil, season, etc. *Beer* is produced by fermenting maltose, the sugar that is formed from starch during the germination of barley (see p. 76). The yeast used is a 'domesticated' yeast that is carefully preserved for year after year by the brewer.* *Spirits* contain more alcohol than beer† or wine and are made by distilling various fermented sugars; thus whisky† comes from fermented barley or potatoes, brandy† from wine, and rum† from fermented sugar-cane molasses (treacle).

In making bread, a little yeast (baker's yeast, a special variety) is added to the dough,* which is then warmed. The carbon dioxide produced during fermentation forms large bubbles in the dough and 'lightens' the bread. The sugar used by the yeast comes from the hydrolysis of starch in the flour; the alcohol that is formed evaporates during baking.

A SMALL-SCALE EXPERIMENT TO SHOW ALCOHOLIC FERMENTATION

Take 1·5 gm. of dried yeast in a mortar, and add to it, a little at a time, 15 c.cm. of a 20 per cent. solution of glucose, grinding thoroughly after each addition. Then completely fill a 6-in. by $\frac{1}{2}$-in. test-tube with the mixture and carefully push this tube up into an inverted 6-in. by 1-in. flat-bottomed tube (a specimen-tube) until the flat bottom of this tube closes the mouth of the test-tube,

as shown in Fig. 32A. The apparatus can then be inverted without any solution escaping from the test-tube or any air entering (see Fig. 32B). (A small air-bubble, however, may be neglected.) The apparatus is then put in the 37° C. water-bath and, after a time, bubbles of gas begin to rise to the top of the test-tube, displacing the liquid into the outer tube (see Fig. 32C). After about half an

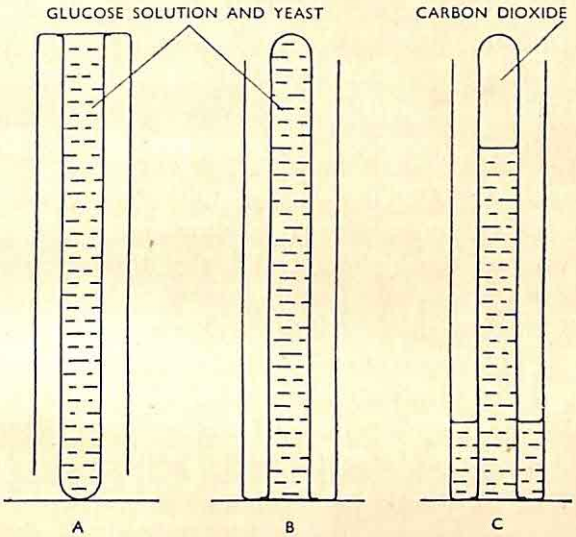


FIG. 32. Fermentation of glucose.

hour the test-tube becomes full of gas. When this stage is reached, remove the apparatus from the water-bath, and, keeping the mouth of the test-tube closed by the bottom of the specimen-tube, pour off the displaced liquid into another test-tube and set it aside to test for alcohol later. Into the outer specimen-tube, pour about 15 c.cm. of sodium hydroxide solution (2N) and then rock the inner tube gently. Notice that the alkali rises inside the test-tube as the carbon dioxide is absorbed, and continues rising until it fills the test-tube, showing that the tube was full of carbon dioxide.

To about 10 c.cm. of liquid from the test-tube in the last experiment add about 10 drops of iodine solution and mix by shaking.

The brown colour of the iodine disappears. Warm gently and notice the characteristic smell of *iodoform*, showing the presence of *ethyl alcohol.*

VINEGAR AND ACETIC ACID

Fermented sugar solutions turn 'sour' when they are exposed to the air, owing to the formation of *acetic acid.* This is formed by another fermentation process brought about by 'acetic bacteria', the spores of which are always floating about in the air. The ethyl alcohol is oxidized to acetic acid:

$$C_2H_5OH + O_2 \longrightarrow CH_3COOH + H_2O$$

Acetic acid—CH_3COOH—is one of the simplest members of another important series of organic compounds—the *fatty acids*—so called because the higher members of the series are present, combined with *glycerine*, in many fats. Acetic acid is present in *vinegar*,† formed by allowing fermented sugar solutions to ferment still further; but concentrated acetic acid is prepared from the watery liquid that distils off when wood is heated in closed retorts (see p. 55). The *acetates*† are salts of acetic acid.

Two important organic acids (but not belonging to the fatty acid series) are *citric acid*† and *tartaric acid*,† both of which are crystalline solids. Citric acid is present in the juice of citrus fruits, e.g. lime, lemon, orange, and has an agreeable sour taste. It is used in making bottled fruit drinks: Tartaric acid is a by-product from wine-making. It is mixed with sodium bicarbonate to make *baking-powder.*† The acid and alkali do not react until the mixture is wetted, when the tartaric acid liberates carbon dioxide from the sodium bicarbonate.

'ORGANIC SALTS'—ESTERS†

The alcohols, although they have no alkaline properties, are rather like the bases we have studied in inorganic chemistry. Just as *acid+base* $\longrightarrow$ *salt+water* in inorganic chemistry, so *fatty acid+alcohol* $\longrightarrow$ *'organic salt'+water* in organic chemistry.

Such 'organic salts' are called *esters*. Most esters are pleasant-smelling liquids and some are used to give an artificial fruity flavour to food and drinks. A simple example of *ester formation* is the reaction between acetic acid and ethyl alcohol to form *ethyl acetate*† and water:

$$CH_3COOH + C_2H_5OH \longrightarrow CH_3COOC_2H_5 + H_2O$$

Mix together in a test-tube about 1 c.cm. each of acetic acid (concentrated), ethyl alcohol, and concentrated sulphuric acid (to remove the water from the reaction), and warm the mixture gently. Notice the sweet 'fruity' smell of ethyl acetate.

OILS AND FATS

Many plants store *oil* as a reserve food material instead of starch, e.g. coconut-oil, palm-oil, soya-bean-oil, olive-oil,† etc. Fishes and other 'cold-blooded' sea animals also produce oils, e.g. cod-liver-oil,† while 'warm-blooded' animals store similar substances with higher melting-points, which are solids at ordinary atmospheric temperatures. These solid oils we call *fats*. (Coconut-oil and palm-oil are *liquid oils* in the tropics but they become *solid fats* in colder countries.)

Such oils and fats are *not* hydrocarbons like petroleum, but they contain the elements *carbon*, *hydrogen*, and *oxygen*. They are *not*, however, carbohydrates, since the hydrogen and oxygen are not present in the same proportions as in water. There is in fats much less oxygen in proportion to hydrogen than in carbohydrates; hence fats have a higher calorific value and require more oxidation than do carbohydrates. Fats are 'organic salts', or esters, being compounds (*glycerides*†) of the 'organic base' *glycerine* (or *glycerol*,† since it is an alcoh*ol*) and *fatty acids* of high molecular weight. One molecule of glycerine combines with three molecules of a fatty acid (since glycerine contains three (OH) groups). For example, olive-oil is mainly composed of the glyceride of *oleic acid*;† palm-oil is mainly the glyceride of *palmitic acid*;† mutton-fat† contains the glyceride of *stearic acid*,† and butter is mainly the glyceride of *butyric acid*.†

When a drop of vegetable oil is placed on a piece of filter-paper, a permanent translucent *grease-spot* is formed. This grease-spot, unlike the translucent spot produced by water or any liquid that evaporates, does not disappear on standing, and it cannot be washed away with water. This is a useful test for fats when present in fairly large quantities.

(i) Place one drop of coconut-oil on a strip of filter-paper and notice that a permanent translucent grease-spot is formed. (Keep this strip of filter-paper for Expt. (v).)

(ii) Place single drops of (*a*) water, (*b*) alcohol, (*c*) petrol, (*d*) glycerine, on strips of filter-paper. Notice that although a translucent spot is first formed in each case, the water, alcohol, and petrol soon evaporate and leave no trace behind. The glycerine, however, does not evaporate. Wash the fourth (glycerine) paper under the tap and then pin it up to dry. Notice that no translucent spot is left after drying the paper.

(iii) If the substance under test contains very little fat, we can make use of the fact that oils and fats are soluble in petrol. Take about 1 gm. of crushed soya-bean or ground-nut in a *dry* test-tube and add about 5 c.cm. of petrol (away from flames). Close the mouth of the tube with the thumb and shake vigorously for several minutes. Then allow the solid to settle. With a narrow glass tube, remove a little petrol and allow it to fall—a drop at a time—on a strip of filter-paper, allowing one drop to evaporate before adding the next drop. Notice that a permanent grease-spot is formed as the petrol evaporates. (Keep this filter-paper for Expt. (v).)

(iv) If the substance contains much fat, it will form a grease-spot if the crushed material is placed on a filter-paper and gently warmed. Try this test with an oil-yielding seed, e.g. sunflower, or castor-oil seed.

(v) Take the strips of filter-paper from Expts. (i) and (iii), place them in a test-tube, and add a few c.cm. of Sudan III solution. Shake and then allow to stand for several minutes. Then remove the strips and wash with water, under the tap. Notice that the

grease-spots are now stained a bright orange-red. This staining test enables us to test for fat in substances that do not give the grease-spot test readily.

SOAPS

There is a whole series of *fatty acids* corresponding to the paraffin series of hydrocarbons, one of the simplest members being acetic acid. *Stearic* and *palmitic* acids are higher members of this series, and *oleic* acid is closely related. When fatty acids are neutralized by alkalis, salts are formed, e.g. when acetic acid is neutralized with sodium hydroxide, sodium acetate† is formed:

$$CH_3COOH + NaOH = CH_3COONa + H_2O$$

Similar salts are formed when any fatty acid is neutralized by a base, but the sodium and potassium salts of the higher fatty acids have special properties and are called *soaps*. The *sodium salts of the higher fatty acids are* '*hard soaps*', while the corresponding *potassium* compounds are '*soft soaps*'.

Soaps are not made commercially by direct neutralization of fatty acids with alkalis, but by boiling fats with caustic alkalis, forming a mixture of soap and glycerine, i.e. *fat+alkali = soap+glycerine*. Fats or oils are mixed with caustic soda and the mixture is heated, either by steam-coils or by blowing steam through the liquid. After the mixture has been heated for some time, all the fat is converted into soap. At this point salt is added. The soap, being insoluble in salt solution, separates out as a solid on the surface, the glycerine remaining behind in solution. The soap is purified and is then pressed into moulds where it cools and hardens. The glycerine, being a very valuable by-product, is recovered by distillation. It is used mainly for making explosives and for medicinal purposes.

Only sodium and potassium soaps are soluble in water. The stearates, oleates, and palmitates of other metals are all *insoluble soaps*. Hence, when soap is added to *hard water* (containing calcium and magnesium salts) no lather is produced at first, but only

a surface layer (or scum*) consisting of the calcium and magnesium soaps, which are insoluble in water. For example:

Sodium stearate† + *calcium sulphate* = *calcium stearate*† + *sodium sulphate*
(*soluble soap*) (*insoluble soap*)

Hence the amount of soap required to produce a permanent lather measures the *hardness* of a sample of water (see Book Two, Chap. I).

TO PREPARE SOME HARD SOAP

(*Demonstration.*) Take a 250 c.cm. flask fitted with a cork carrying a long, wide, glass tube (4–5 ft. long). Put 25 c.cm. of coconut-oil in the flask, and add 100 c.cm. of alcoholic sodium hydroxide solution (4 gm. NaOH in 100 c.cm. of alcohol or methylated spirit). Replace the cork and heat over a water-bath. By using alcoholic caustic soda, we obtain a reaction very quickly, since the oil dissolves in the alcohol where it can react with the alkali. Any alcohol that boils away condenses in the long glass tube and is returned to the flask. After 10 minutes, quickly pour out a few drops of the clear solution into some water in a test-tube and replace the flask on the water-bath. If drops of oil separate in the water in the test-tube, the reaction is not yet complete. If a clear solution is obtained, all the oil has been converted to soluble soap. When the reaction is complete, pour the hot solution into a basin and allow it to cool. The alcohol soon evaporates, leaving a solid mass of soap (mixed with glycerine).

(*Class Experiments.*) (*a*) Take a small, stoppered bottle (about 100 c.cm. or 4 fluid oz.) and put in it 25 c.cm. of coconut-oil. Take 20 c.cm. of 20 per cent. sodium hydroxide solution in a measuring cylinder, and add about 1 c.cm. of this to the oil. Shake the bottle *vigorously* for 2 minutes (so as to break up the oil into tiny droplets, thus increasing the surface area of oil and alkali in contact with each other). Then add 5 c.cm. of caustic soda and shake for 1 minute, and so on until all the alkali has been added. Then pour out about half this mixture into a 100 c.cm. beaker and

allow it to stand while you do the following experiments with the remainder of the soap.

The success of this method depends upon adding the alkali very slowly at first and shaking very vigorously. It also works better with some samples of coconut oil than with others. The following experiment gives an alternative method.

(*b*) Weigh out 4 gm. of solid sodium hydroxide in an evaporating-dish and dissolve it in a mixture of 10 c.cm. of water and 5 c.cm. of alcohol. Then add a mixture of 25 c.cm. of coconut oil and 5 c.cm. of alcohol slowly, stirring all the time with a glass rod. Heat on a boiling water-bath with constant stirring until the mixture becomes stiff (10–15 minutes), showing that the reaction is complete.

EXPERIMENTS WITH SOAP

(i) Half-fill the stoppered bottle from Expt. (*a*) above with warm water and shake vigorously to dissolve the soap. Notice the characteristic lather—a mass of air-bubbles, each bubble surrounded by a film of soap. Keep this solution for the following experiments.

(ii) Put about 25 c.cm. of saturated brine into a 100 c.cm. beaker and pour into it an equal volume of the soap solution. Notice that the soap is 'salted out' and rises to the surface.

(iii) Take about 10 c.cm. of the warm soap solution in a test-tube, add to it about 1 c.cm. of concentrated hydrochloric acid and shake. The fatty acids are set free, forming a milky suspension. On standing, the fatty acids rise to the surface and form a separate oily layer that solidifies on cooling.

(iv) *Emulsions.* Take three test-tubes, A, B, and C. Put 10 drops of coconut oil in each tube. To A add about 10 c.cm. of water; to B add 10 c.cm. of water and a few drops of sodium hydroxide solution; to C add 5 c.cm. of water and 5 c.cm. of your soap solution. Shake all three tubes vigorously and then allow them to stand in the test-tube rack. In each case vigorous shaking breaks up the oil into very fine drops that become suspended in the liquid

to form an *emulsion*.† On standing, however, the oil-drops in A soon run together and rise to the surface to form a separate layer, i.e. the emulsion is unstable. In B the emulsion is permanent, because some alkali combines with some oil to form a little soap, which forms a protective film round each tiny oil-drop. In C the emulsion is also stable, for the same reason.

In the same way, when greasy or oily articles are washed with soap, the soap solution '*emulsifies*' the grease, and this emulsion can easily be washed away with water.

(v) (*a*) Put a little lamp-black (finely-divided carbon) in a test-tube and add about 10 c.cm. of water. Shake vigorously. Filter half the suspension and allow the remainder to stand. The filtrate is colourless, all the solid particles being retained by the filter. On standing, the solid particles gradually settle out from the suspension.

(*b*) Put about the same quantity of lamp-black in another test-tube and add 5 c.cm. of water and 5 c.cm. of the soap solution. Shake vigorously. Filter half the mixture, as before, and notice that the filtrate is black, each separate particle of lamp-black having been surrounded by a film of soap and thus enabled to pass through the filter. Allow the remainder of the mixture to stand—the lamp-black remains suspended. This experiment explains how soap removes dirt, each particle of dirt being surrounded by a film of soap so that it is readily washed away with water.

PROTEINS

Proteins are the third great class of organic compounds present in plants and animals along with carbohydrates and fats. Seeds are always rich in proteins, and small quantities of proteins are found in roots, and stems, and leaves. The bodies of animals always contain a larger proportion of protein than is found in plants. These animal proteins are all obtained, in the first place, from plant proteins. Plants synthesize their proteins from simple raw materials like sugars, nitrates, and ammonium compounds.

Proteins always contain carbon, hydrogen, and oxygen *together*

with nitrogen. Nearly all the proteins also contain *sulphur*, while many of them contain *phosphorus* as well. Proteins are some of the most complex of all chemical compounds, their molecules being very large and containing enormous numbers of atoms; for example, the *haemoglobin* molecule is said to have the formula $C_{3032}H_{4816}O_{872}N_{780}S_8Fe_4$. In a General Science Course of this kind we cannot attempt more than an outline of the chemistry of the proteins, but because of their great everyday importance as essential constituents of all living things and as food-materials we cannot leave them out.

AMINO-ACIDS

When proteins are boiled with acids for a long time, they are split up by hydrolysis into simpler compounds called *amino-acids*;† in fact, a protein molecule appears to consist of a long chain of amino-acids. There may be 50 to 500 links in this chain, and each link is an amino-acid molecule. These amino-acid 'links' can be 'unlocked' by heating with acids or by the action of *protein-splitting enzymes* (just as the polysaccharides and disaccharides can be hydrolysed into monosaccharides). Or, to put it another way, the amino-acids are the units, or 'bricks', from which proteins are built.

About 20 of these amino-acids are found in natural proteins, and one protein differs from another in the number and arrangement of its constituent amino-acids.

The simplest amino-acid is *glycine*.† It has a sweet taste and was first prepared by boiling glue or gelatine with acids. The chemical name for glycine is *amino-acetic acid*,† as it is derived* from acetic acid, CH_3COOH, one hydrogen atom of the (CH_3) group being replaced by the *amino*† group (NH_2), giving $CH_2(NH_2)COOH$. All the amino-acids can be regarded similarly as being derived from fatty acids and ammonia.

This has an important bearing on the use of proteins as foods. The proteins used as food by animals are very seldom the same as those in the animal's body-tissues, and digestion of proteins serves

to split up the food-proteins into their constituent amino-acids before they enter the blood-stream. The body-cells then select from these amino-acids the ones that can be built up into the kind of body-protein they use for growth and repair. The rejected amino-acids are used as fuel, to supply energy.

During digestion, the very complex proteins are broken down by stages:

PROTEINS ⟶ PROTEOSES† ⟶ PEPTONES† ⟶ POLYPEPTIDES† ⟶ AMINO-ACIDS

We shall learn more about these changes later.

All proteins are not of equal food-value, because they do not all contain all the essential amino-acids in suitable proportions. A protein that contains all the essential amino-acids in the proportions necessary to support human life is called a 'complete' (or 'first-class') protein, or it is said to be a protein of good biological value. A protein that does not contain all the essential amino-acids, or which contains them in unsuitable proportions, is called an 'incomplete' (or 'second-class') protein, or it is said to be a protein of poor biological value. All animal proteins (except gelatine) are 'complete' proteins, while plant proteins are all 'incomplete'. Examples of animal protein are lean meat (beef, mutton, fish, and birds), eggs, milk, and cheese. These are all expensive foods. Examples of foodstuffs supplying vegetable proteins are the various kinds of legumes (pod-bearing plants—peas, beans, pulses, lentils, soya-beans). These plant proteins are cheaper than animal proteins.

TESTS FOR PROTEINS

Nearly all the tests for proteins depend upon reactions given by the constituent amino-acids; hence *all* proteins do *not* give *all* the colour reactions, so that it is necessary to carry out several tests before the presence of protein is verified. Try the following tests on a solution of egg-white.

BIURET REACTION†—THE PROTEIN–COPPER TEST

This test is given by all *proteins* (in which the individual amino-

acids are linked together) but *not by the individual amino-acids*. Take 2–3 c.cm. of the test-solution in a test-tube and add an equal volume of sodium hydroxide solution (2N=8 per cent.). Then add *two drops* of 1 per cent. copper sulphate solution, and shake gently to mix. Proteins give a *violet* colour at once, without heating. (The less complex proteoses and peptones give a reddish or pink colour.) In doubtful cases a control tube should be used, with 2–3 c.cm. of water, an equal volume of sodium hydroxide solution, and two drops of 1 per cent. copper sulphate solution, comparing the colour of this liquid with that of the test solution.

NITRIC ACID REACTION FOR PROTEINS

This is given by most amino-acids (either the individual amino-acids or the amino-acids when present in combination in a protein). Take 2–3 c.cm. of the test-solution and add about 1 c.cm. of concentrated nitric acid. If one of these amino-acids is present, a white precipitate is formed that turns *yellow* and partly dissolves on warming. *Cool* under the tap and then add excess of ammonia (*care!*). The yellow colour turns to orange. You should now be able to explain why nitric acid, when accidentally spilt on the skin, produces a yellow stain.

VIRUS PROTEINS

In earlier lessons we made a clear distinction between living and non-living things, but there have recently been discovered certain pure chemicals, the *virus proteins*,† which show living properties. There are some diseases of plants and animals that resemble bacterial diseases because a small quantity of material from an infected plant or animal can infect a healthy individual. But when this infectious material is examined under the microscope it shows *no cells*, so the disease cannot be caused by bacteria or other one-celled organisms. One of these diseases—tobacco mosaic disease†—produces a characteristic pattern on the leaves of infected tobacco plants. If the juice from the infected leaves is squeezed

out and passed through the finest filters (which would retain any *cells*) the filtrate will still infect healthy leaves. Such a disease-producing agent is called a *virus*, or a *filterable virus*.

Tobacco mosiac virus has been proved to be a *protein*, and solid crystals of this protein (a pure chemical) have been obtained from the infectious plant juice. Crystals of this virus-protein can be stored, like any other pure chemical, without showing signs of life. But if a crystal of the virus-protein is put on a healthy tobacco leaf it produces mosiac disease, which spreads over the whole leaf, showing that *the virus protein is making more of itself*, i.e. *it grows and reproduces itself like all living things.*

Virus proteins, therefore, seem to bridge the gap between living and non-living things, and it is thought that such very complex chemical compounds may have given rise to more complex living *cells*, which are the structural units of most living things. But virus proteins are non-living (when stored in a test-tube, for example) and only 'come to life' (living as *parasites*) when they come into contact with suitable living cells. Human diseases caused by viruses are discussed in Book III A.

(See *General Science Workbook*, Book Two, pp. 144–8 for learning exercises on this chapter.)

SECTION II
PHYSIOLOGY

CHAPTER IV
THE STRUCTURE AND FUNCTIONS OF LIVING MATTER

CELLS

In 1665, Robert Hooke, examining thin slices of bottle-cork (cork-oak† bark) under the newly invented microscope, noticed that the cork, when magnified, looked very much like a honeycomb, being built up of a large number of closely-fitting little boxes, to which he gave the name *cells*. The cork-cells examined by Hooke were actually *dead cells*, and it was not until nearly 200 years later that there was any clear idea of the true nature of cells and that *the most important part of a cell is its living contents and not its dead cell-walls.*

All living matter, whether plant or animal, is alike in being built up of these tiny units called *cells*. Some of the simplest living things consist of one single cell, e.g. Amoeba, but a 'higher' plant or a 'higher' animal consists of millions of these cell-units. All cells, however, are not alike, for they have different *functions* to perform. For example, the cells in the skin of a leaf are different from those in the middle of a leaf; the cells of animal muscle are different from nerve-cells, and so on. Collections of similar cells, all doing similar work, are called *tissues*. In 'higher' plants, for example, there is one type of cell that forms the *strengthening tissue* in the stem, and another kind that forms the *conducting tissue* through which manufactured food-materials travel from the leaves to other parts of the plant. In 'higher' animals there is skin-tissue, nerve-tissue, muscle-tissue, bone-tissue, gland-tissue, etc., all carrying out different functions and built up of different kinds of cells.

Cells, therefore, are the tiny living units from which plants and animals are built up.

CELL STRUCTURE

A young *plant cell* consists of a tiny 'box' with a thin *cell-wall* composed of the carbohydrate *cellulose*. This non-living cell-wall encloses the actual living matter—*protoplasm*[1]—which is the basis of all *life*. Protoplasm is a colourless, almost transparent, jelly-like substance, somewhat like the 'white' of a raw egg, and of very complex chemical composition. The protoplasm of almost every cell contains a denser, rounded portion called the *nucleus*, which seems to control the working of the cell, e.g. when an Amoeba is cut in two, the part containing the nucleus lives and grows normally while the part without a nucleus dies. Fine threads of protoplasm pass through tiny holes in the cell-wall and join up the protoplasm of neighbouring cells, so that the protoplasm is continuous from end to end of the plant. In very young plant-cells, e.g. those at the growing point of the stem or root, the protoplasm fills the whole of the cell, but as the cell grows bigger, *cell-spaces* appear in the protoplasm, and these spaces are filled with *cell-sap*, a watery solution of sugars, proteins, mineral salts, and dissolved gases. As the cell grows still larger, the protoplasm merely forms a living lining round the inside of the cell-wall, and the middle of the cell is filled with sap. In still older cells, e.g. the *heart-wood* of an old tree, the protoplasm disappears altogether and the cells are no longer living (although this dead tissue is still useful to the tree in providing strength).

An animal-cell also consists of protoplasm with a nucleus, but it is surrounded by a very thin, *living*, cell-membrane (composed of a firmer layer of *protoplasm*) and is not enclosed by a *dead cell-wall* (composed of *cellulose*) as in plant-cells. This absence of a non-living external layer of cellulose is one of the most characteris-

[1] The term *protoplasm* was invented in 1861 to denote the 'universal living substance' of which organisms are made. It is now being replaced by the more precise term *cytoplasm*,† or cell-material, which denotes all the protoplasm of the cell except the nucleus.

tic differences between the cells of plants and animals, and for this reason plant-cells are more definite in shape and show a clearer outline when examined under the microscope. Another characteristic difference between plant- and animal-cells is that fully-grown plant-cells usually have cell-spaces (filled with cell-sap) occupying a large part of the cell, while animal-cells never have these large cell-spaces but are nearly always full of protoplasm.

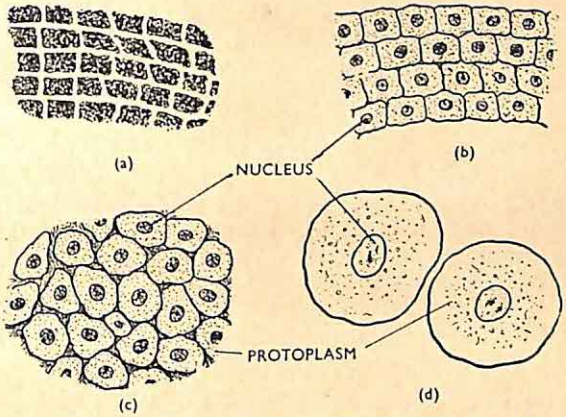


FIG. 33. Cells (highly magnified). (*a*) Hooke's drawing of dead cork cells. (*b*) Plant-cells from young root. (*c*) Frog skin-cells. (*d*) Cells from inside human cheek.

EXAMINATION OF CELLS

Plant-cells. (i) Peel off, with forceps, a small piece of the transparent outer skin from the scale-leaf of a bulb, e.g. onion or lily, or from a green leaf, e.g. Allamanda. Mount it in a drop of water, under a cover-glass, and examine it with the microscope or microprojector. Notice (*a*) the *cell-wall*, and (*b*) the *nucleus* in the middle of the protoplasm that fills the cell.

Animal-cells. (ii) Examine some animal-cells by scraping the inside of your cheek with a clean finger-nail and mounting the scraping in a drop of saliva under a cover-glass. Notice that these cells, unlike plant-cells, have *no cell-wall.* This is one of the main differences between plant- and animal-cells—*plant-cells have a*

non-living cellulose cell-wall, while animal-cells have no such non-living cell-wall but a living membrane of protoplasm that is rather more solid than the remainder.

(iii) Examine under the microscope a little of the soft pulp from inside a tomato. Notice the rounded cells.

(iv) Mount a young leaf of moss or water-weed in a drop of water under a cover-glass and examine with the microscope. Under the high power, a continuous current of protoplasm can usually be seen, moving slowly round the walls of the cell and carrying the green *plastids* with it. (This movement is usually most rapid in the centre of the leaf, in cells either in or adjoining the mid-rib.)

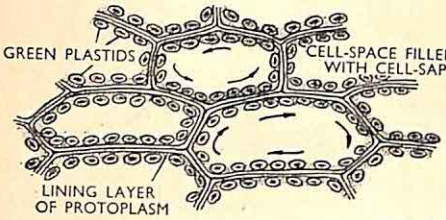


FIG. 34. Movement of protoplasm in leaf-cells of water-weed (highly magnified).

(v) Examine a few threads of Spirogyra under the microscope and notice the cell structure. Mount some more Spirogyra in a drop of 10 per cent. sugar solution and examine under the microscope. The solution diffuses through the cell-wall but cannot pass through the lining layer of protoplasm, which is thus forced away from the cell-wall and is more easily seen.

THE FUNCTIONS OF LIVING PROTOPLASM

A one-celled animal like Amoeba, or a single cell of a simple green plant like Spirogyra, can carry out all the following essential functions: (*a*) *nutrition* (or feeding), (*b*) *respiration*, (*c*) *excretion* (or getting rid of waste products), (*d*) *moving*, (*e*) *feeling* (or response to external stimuli), (*f*) *growing*, and (*g*) *reproduction*. *These are the fundamental functions of living protoplasm.* Since all living things are built up of one or more cells of protoplasm, these are also the fundamental functions of *all living things*. In the 'lowest', one-celled, plants and animals one single cell can perform all these functions for itself. In the 'higher' plants and animals, consisting

of many cells, there is *specialization*, and some cells are specially good at performing one special function. Thus some cells (muscle-cells) are specially good at contraction, some (nerve-cells) are specially good at feeling, while others (wood- and bone-cells) are specially strengthened, and so on.

In the next five chapters we shall study these functions of living protoplasm in turn, examining the different ways in which different living things carry out the same functions. This branch of science is called *physiology*: the study of the way in which living things work.

NUTRITION AND METABOLISM IN LIVING THINGS

Living protoplasm is in a constant state of change; in it there are always going on a great many chemical and physical processes that together make up what we call *metabolism*†—a word that we can translate as 'changing-ness'. These processes are most complicated and we do not yet know much about them. But this much is clear: some of them are *anabolic*† or *building-up processes* by which are made the large molecules of protoplasm and the body-structure; others are *katabolic*† or *breaking-down processes* during which some of these previously formed molecules are partially decomposed. A special and universal kind of katabolic reaction is *respiration*, during which sugars are oxidized. This process, besides leading to a loss of sugar, liberates energy, which is used in movement and to drive the reactions concerned in growth and repair.

Some (but by no means all) of the reactions going on in living cells are not stopped if the cells are damaged or even ground up into fine particles; some can even be imitated by the chemist in the laboratory. But the special feature of *metabolism* is that *there is a balance between the various processes*. Protoplasm is unstable; it is constantly being built up and broken down, and as long as this balance is maintained the protoplasm remains *alive*; if the balance is upset, the organism becomes unhealthy; if the balance is destroyed, the organism dies.

Chapter V

NUTRITION IN PLANTS AND ANIMALS

A mature* organism (such as a man that has stopped growing) is constantly respiring. It is therefore losing some of its carbon in the form of gas, carbon dioxide. At the same time it is also losing some of its carbon and nitrogen compounds by excretion (e.g. through the kidneys in Man). Hence the organism continually needs food to replace these losses. Very often it is also laying down storage materials, which make a further demand on its food and (if the organism is *growing*) still more food is required for the formation and upkeep of new tissues. Animals are much more active than plants and a greater proportion of their food is used in respiration. Plants, on the other hand, use up a great deal of carbohydrate in making new skeletons of cellulose and wood, and they often store excess food in their seeds, underground stems, and roots.

The essential living material of both plants and animals is protoplasm. This always requires the same classes of food substances—carbohydrates, fats, proteins, mineral salts, and water. The food of the organism *as a whole*, however, is not the same for plants and animals.

Animals get their organic foods ready-made, e.g. they eat meat, plants, etc. Green plants make their own foods for themselves from simple raw materials—from the carbon dioxide of the air, together with the water and mineral salts that they get from the soil.

NUTRITION IN GREEN PLANTS

The green plant builds up carbohydrates from carbon dioxide and water, the necessary energy being obtained from sunlight. This process goes on only in certain green parts of the protoplasm, called *plastids*. These are usually small, dense, oval masses of protoplasm found mainly in the cells of green leaves and in the

outer cells of green stems. *Colourless* plastids also occur in root cells and in underground storage stems. These can make and store starch but they cannot carry out photosynthesis even when illuminated. For this process the green colouring substance *chlorophyll* is essential. Yet this substance, when extracted from the green plastids, cannot bring about photosynthesis. Neither can undamaged plastids, isolated from the plant cell. *The process only goes on in green illuminated plastids supplied with carbon dioxide and still contained in an undamaged cell.*

LEAF STRUCTURE

The leaves of different plants are of different shapes and sizes, but, in general, nearly every leaf exposes to the light and air two broad, flat, green surfaces.

Each leaf has a 'skeleton' of veins; these are continuations of the *conducting* or *vascular* bundles of the stem and its branches. Each vein is made up of two threads, the *xylem*† (wood) on the upper side and the *phloem*† (bast) on the lower side. If a thin cross-section of a leaf is examined under the microscope, it is seen to have a leaf-skin on both surfaces consisting of a single layer of flat, transparent, non-green cells (see Fig. 35). In the under-surface (and sometimes in both surfaces, as in monocotyledons) are very small pores, the stomates. Each pore is bounded by two *guard cells*, which can change their shape and so regulate the size of the stomatal pore. The surface appearance of stomates is shown in Fig. 67, Book Two.

Just beneath the upper leaf-skin is a layer of long cells, called *palisade cells*,† packed closely together perpendicular to the leaf-surface. These palisade cells contain most of the green plastids of the leaf, and since this part of the leaf gets most light, it is in these cells that most carbohydrate is synthesized. The cells in the lower part of the leaf, below the palisade cells, are usually loosely packed, forming a *spongy tissue*, with plenty of *air-spaces* between the cells. These air-spaces communicate with the narrower ones between the palisade cells and also with the atmosphere through

the stomates. It is through the stomatal pores that exchange of gases and loss of water-vapour take place. Usually the pores are open in the day-time and closed at night, but they may close at any time if the plant loses more water than it takes up, i.e. if it *wilts.*

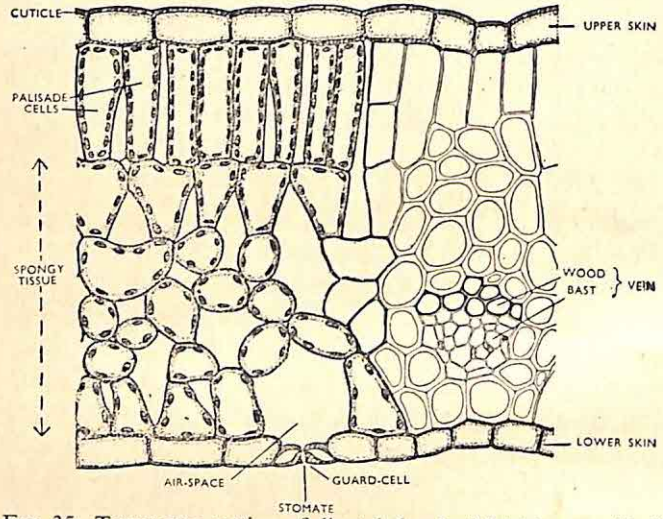


FIG. 35. Transverse section of dicotyledon leaf (highly magnified).

Complete closure of the stomates reduces the loss of water from a leaf, but it means also that photosynthesis must stop.

METABOLISM IN THE LEAF

When light falls on a green leaf, sugars are made in the green plastids very much faster than they are removed from them. The excess sugar is stored temporarily in the plastids during the day, as starch, or (in monocotyledons) it is stored in the cell as sucrose (cane sugar). Of the remainder, some is used in the leaf-cells for respiration and in making amino-acids, and some travels away from the leaf along the vascular bundles. The amino-acids do not accumulate in the leaf-cells but move to other parts of the plants, where they are used in synthesizing proteins.

During the hours of darkness, when photosynthesis stops, the movement of sugar away from the leaf continues and the tem-

porarily stored starch disappears from the plastids. It is converted into sugar by hydrolysis, catalysed by the enzyme *diastase*. Thus, by morning, the green leaf is normally free from starch again.

The water used in photosynthesis (as well as that lost in transpiration) reaches the leaf through the *xylem* strands of the veins. This is a tissue largely made up of long-dead cells with hard woody walls. It is continuous with the same tissue in the stem and root. Dissolved in the water moving in these tubular cells are mineral salts that have entered from the soil. Of these, *nitrates* are used in the leaf for making amino-acids from the sugars. Phosphates and sulphates from the soil provide two other elements (phosphorus and sulphur) that form part of some of these amino-acids. Other mineral substances also play an important part in metabolism, though not only in mature leaves (see Book Two, Chap. VI).

We have earlier carried out simple experiments that support some of the statements made above. It is not easy to show exactly how sugars and organic nitrogen compounds are removed from the leaf. We know, however, that they do *not* travel in the xylem, i.e. in the transpiration stream, but in the other conducting strands of the vascular-bundles, the *phloem*. The conducting cells of the phloem are not dead woody tubes, but long, thin-walled, living cells connected end to end. It is along these that organic compounds, like sugar, move up and down the plant. One proof of this is given by *ringing* experiments. In a *woody* stem or tree-trunk (see Fig. 36), the phloem is confined to a narrow cylinder outside the wood—it forms a thin layer just inside the *bark*. If a complete ring of bark and phloem is removed (as when a tree is *ring-barked*),[1]

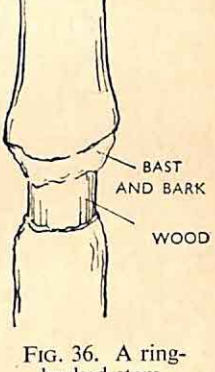


FIG. 36. A ring-barked stem.

[1] In everyday conversation the term *bark* usually denotes everything outside the wood, thus including both living *bast* (or phloem) and dead bark. To the botanist, however, the term *bark* is restricted to the protective, corky tissue of dead cells on the outside of older stems and roots of woody plants, e.g. on tree trunks, and does *not* include the underlying bast or phloem (see Fig. 96).

water is still able to pass up to the leaves in the undamaged wood, but sugars and amino-acids can no longer pass down to the roots. Careful observation shows that: (*a*) Soon after 'ringing', the phloem tissues just above the ring become richer in sugars, whose downward passage is now blocked. (*b*) The roots, thus deprived* of food, die as soon as their reserve of food has been used up.

. The appearance of the conducting tissues in cross-sections of soft and woody stems is shown in Figs. 94 and 96.

The organic substances made in leaves may be stored temporarily (e.g. as starch), but they all pass eventually to other parts of the plant. You will often find it stated that they pass to the parts that '*need*' them—for respiration, growth, or storage. This is not a very sound explanation unless we are quite clear about what we mean by the 'needs' of a plant. What we do *know* (as the result of experiments) is that the sugars and amino-acids each move in the phloem from any part of the plant in which they are in high concentration to another part where they are in low concentration, i.e. to one in which they are being used up or changed to storage material. Thus sugars pass to all cells where respiration is going on, but they move in the greatest quantities to those cells that are (*a*) growing, or (*b*) storing starch (or oil). The movement is therefore away from the leaves towards the stem- and root-tips, to developing seeds and fruits, and to underground storage organs.

STORAGE ORGANS IN PLANTS

We have seen that in animals, food is stored mainly in the liver and in fatty tissues (e.g. under the skin). Plants very often develop specialized storage organs that vary from one species to another. In seeds, starch, fats, and proteins may be stored, either in the swollen seed-leaves (cotyledons) or in a special tissue called the *endosperm*.† Sometimes roots grow very thick (as in carrots) and store large quantities of starch or sugar in their cells. More often it is *underground stems* that are modified for storage. Potato *tubers* are greatly enlarged stem-structures bearing only tiny scale-leaves and buds ('eyes'); the cells are very rich in starch. *Corms*,

like those of Gloriosa and Colocasia, are swollen underground stem-bases bearing scale-leaves. *Bulbs* (e.g. lilies and onions) are really enlarged stem-*buds* in which sugars are stored mainly in the leaf-bases and in special colourless, fleshy scale-leaves. *Rhizomes* (or rootstocks), like those of Canna and Ginger, are swollen, horizontal underground stems.

Before the reserve foodstuffs stored in the cells or storage organs can be transported and used, they must be changed into simpler soluble substances that will diffuse through cell membranes and travel in the phloem. Such changes in plants are equivalent to digestive changes in animals, but they take place *inside the cells* and not in a special digestive tube. They are hastened by organic catalysts, or *enzymes*. Thus *amylase*† (or *diastase*) speeds up the hydrolysis of starch to sugar, *proteases*† speed up the reaction between proteins and water to form amino-acids, *lipase*† is concerned with the hydrolysis of oils and fats to glycerine and fatty acids. All these enzymes are made by the cells of the storage tissues. Their action can be studied most easily during the germination of seeds.

CELLULOSE AND LIGNIN

The process of photosynthesis is so rapid in green plants that they always produce carbohydrates faster than they can build them into proteins or respire them. Some of the excess sugar is converted, as we have seen, to starch, but a great deal of sugar is changed, in every growing cell, into another substance, *cellulose*. This does not, like starch, form grains inside plastids (either green or colourless), but is laid down as a thin transparent *cell-wall* on the outside of each cell. Moreover, this substance does not serve as a store of sugar. Once made, cellulose cannot be changed back to sugar by the plant, but remains as a permanent plant skeleton. The cellulose of the xylem cells (which conduct water) is changed into a hard woody substance called *lignin*,† after which these cells die and merely act as dead woody tubes. These dead *lignified cells* serve two purposes—*water conduction* and *strengthening*. In

bushes and trees the greater part of the central cylinder of the stem and root is made of these conducting xylem cells (*sap-wood*), and when these are present in large numbers the inner form non-conducting *heart-wood*, which serves such useful purposes in everyday life.

NUTRITION IN NON-GREEN PLANTS

Members of the two groups of organisms, the *Bacteria* and the *Fungi*, contain no green plastids, hence many bacteria and all fungi are unable to make their own sugars. Like the animals they must get these foods ready-made. Some of them are *parasites*, getting their food directly from other *living* organisms; others are *saprophytes*†—getting their food from the *dead* remains or other products of plants and animals; still others live together in partnership with higher plants and animals without harming them, sometimes even helping them. Thus the cellulose-decomposing bacteria in a herbivorous animal's *caecum* are really saprophytes, but they live inside the animal's intestine and indirectly provide their 'host' with some food. The nitrogen-fixing bacteria of the legume-plants live in nodules on their 'host's' roots and so get their sugars; 'in return' they 'fix' atmospheric nitrogen, some of which becomes available to their 'host'.

Like the green plants, bacteria and fungi get their water and salts by absorption through their cell-walls; through these, also, they take in their organic food. Many of these organisms live in soil, in plant cells, in milk, or in fruit juice, where the food substances are already present in solution. Others *secrete** enzymes that bring about the digestion of solid foods, which can then be absorbed. Thus, as regards nutrition, the non-green plants occupy an intermediate position between the green plants and the animals.

NUTRITION IN MAMMALS

Animals get their organic food—carbohydrates, fats, and proteins—'ready-made', and the original source of all such foods is the green plant. The structure and method of life of any animal depends mainly on its method of feeding. This accounts for the

much greater variety of animal types than of plant types. A green plant is usually surrounded by suitable raw food materials—carbon dioxide in the air and water in the soil; and it takes in these plant nutrients most efficiently by remaining fixed to one spot, spreading out its flattened leaves in the air where they will receive as much light as possible, and absorbing water by means of a much-branched root-system in the soil. But an animal usually has to move about in search of suitable food. The food must then be taken inside the animal's body to be digested and absorbed. The various kinds of animals show many adaptations connected with 'food-capture'. For example, insects have complicated mouth parts adapted for biting or sucking; among the birds we find many different types of beaks and feet; among the mammals we find several different arrangements of cutting-, tearing-, and grinding-teeth in flesh-eaters and plant-eaters. Hence, although nutrition is the same essential function in all living things, the details of the process are more varied and more complicated in animals than in plants.

For convenience, we shall consider animal nutrition in four stages: (i) *ingestion*†—the capture or intake of food, (ii) *digestion*—the process of breaking down complex food materials into simpler, soluble, and diffusible substances that can enter the body-cells, (iii) *absorption*—the intake of soluble and diffusible food-substances by the true interior of the body, i.e. by the body-cells, and (iv) *assimilation*†—the building-up of these food-substances into the chemical compounds that the body 'needs'.

INGESTION

In Man, food is taken in at the mouth with the aid of the teeth and lips. If the solid food is in large lumps, smaller pieces are bitten off by the cutting-teeth (incisors) at the front of the mouth (see Fig. 76). This food is then chewed, or ground up, by the grinding-teeth (molars) at the back of the jaws. While the teeth are tearing and grinding the food, it is wetted and softened by *saliva*, and worked into a round ball by the tongue in readiness for swallowing.

By thorough chewing, the solid food is broken up into small pieces that will offer a large surface area to the action of digestive juices in the food-canal. (N.B.—A liquid can only react with an insoluble solid *at its surface*, and the finer the state of division of a given weight of a solid, the greater is its surface area.)

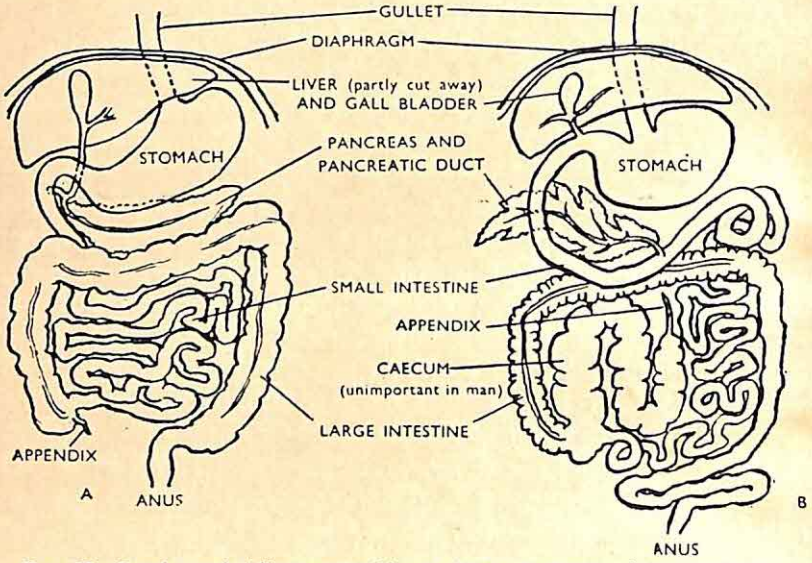


FIG. 37. Food canals (diagrammatic). A. Man. B. Rabbit (after Hatfield).

DIGESTION

Digestion prepares complex foodstuffs for absorption. After ingestion, the food passes through a long tube, the food-canal (or alimentary canal), leading from the mouth to the anus, and the complex carbohydrates, fats, and proteins are broken down into simpler, soluble substances whose molecules are small enough to diffuse through the walls of the food-canal into the true interior of the body. Digestion is brought about by the action of various *enzymes* that are poured into the food-canal by a series of *digestive glands*. These enzymes are *organic catalysts* produced by living cells. There are many chemical changes that take place very slowly

when left to themselves, but which take place rapidly when a small quantity of a suitable *catalyst* is added. Every living cell produces organic catalysts, or enzymes, that enable the cell to carry on its life processes at a suitable speed. Each digestive gland is associated with a different part of the food-canal and produces a digestive juice containing its own special enzymes that act on part of the food as it passes through the food-canal. The enzymes concerned with digestion may be classified as (i) *carbohydrate-splitting enzymes* (*amylases*), (ii) *fat-splitting enzymes* (*lipases*), and (iii) *protein-splitting enzymes* (*proteases*). All these enzymes are destroyed by boiling. If the chemist wants to break down a protein into smaller molecules or to build up a complex organic molecule, he uses high temperatures and pressures, strong acids and alkalis, special solvents, and many other devices. The living organism uses *enzymes* to produce the same reactions at body-temperature, at ordinary atmospheric pressure, and without any solvent other than water. Each enzyme catalyses only one particular chemical change and is usually without effect on other chemical reactions, e.g. there are different enzymes for hydrolysing carbohydrates, fats, and proteins. The enzyme does not seem to enter directly into the chemical change or into the chemical composition of the products, and it is not used up during the reaction, i.e. the quantity of catalyst present at the beginning of the reaction is the same as that remaining at the end of the reaction. In this way, a small amount of enzyme is used over and over again, thus bringing about chemical changes in large quantities of reacting materials.

DIGESTION IN THE MOUTH

In Man, digestion begins in the mouth, where the food is mixed with *saliva*, a slightly alkaline juice produced by the salivary glands. These glands produce 2–3 pints (= 1–1½ litres) of saliva daily. Man has three pairs of salivary glands; (*a*) a pair by the angle of the jaws, just beneath the ears, (*b*) another pair between the two halves of the lower jaw, and (*c*) a third pair beneath the tongue. Saliva softens the food and helps in swallowing, but it is also the

first digestive juice, containing the enzyme *ptyalin* that changes cooked starch into sugar. In practice, the food is not kept in the mouth long enough to change much starch into sugar; hence the digestion of starch only *begins* in the mouth; it is completed later on, in the stomach and in the small intestine. Ptyalin acts on cooked starch, glycogen, and dextrin, converting them into the complex sugar *maltose*, and its activity is increased when common salt (sodium chloride) is present. Ptyalin has no action on complex sugars (e.g. sucrose and maltose), uncooked starch, fats, or proteins. (See pp. 119–120 for experiments illustrating the action of saliva.)

DIGESTION IN THE STOMACH

After being chewed and mixed with saliva, the ball of food is passed down the throat and gullet to the stomach, where it remains for several hours. Here it meets the *gastric juice*, produced by many tiny glands in the lining of the stomach-wall. These glands produce 4–6 pints (= 2–3 litres) of gastric juice daily. This is strongly acid, owing to the presence in it of about 0·2 per cent. of hydrochloric acid; hence it gradually stops the action of the alkaline saliva as the food becomes mixed with gastric juice. In practice, this takes some time, since the last food to be swallowed goes to the *middle* of the stomach and the alkaline saliva is not neutralized by the acid juice until later (see Fig. 38). Hence digestion of starch by saliva continues in the stomach (for about half an hour). The hydrochloric acid in the gastric juice also kills most bacteria and thus prevents food from fermenting in the stomach.

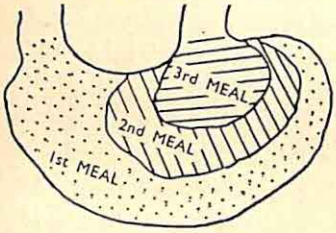


FIG. 38. Section of frozen stomach of rat after three differently coloured meals (after Roaf).

Gastric juice contains two enzymes, (*a*) *pepsin*,† which in acid solution begins the digestion of proteins, splitting them up into simpler, more soluble substances (but without acting on the carbohydrates and fats), and (*b*) *rennin*,† which clots milk so that it is retained in the stomach while pepsin acts on the milk-protein.

In the stomach, food and gastric juice are gradually mixed into a soft liquid mixture and, at intervals, a ring of muscle between the stomach and the small intestine relaxes and the partly-digested food is passed, a little at a time, through an opening called the *pylorus*† into the *duodenum*,† the first part of the small intestine, where it meets more digestive juices. (See pp. 120–2 for experiments illustrating the action of pepsin and rennin.)

DIGESTION IN THE SMALL INTESTINE

When the acid contents of the stomach enter the small intestine they are neutralized by alkaline *bile* (from the liver) and alkaline *pancreatic juice* (from the pancreas). In the first part of the small intestine (the *duodenum*), starches, proteins, and fats are all split up by the action of three enzymes in the pancreatic juice that will be discussed later.

Bile is a greenish, alkaline fluid made by the liver and stored in the *gall-bladder*, 1–2 pints (= 500–1,200 c.cm.) being produced daily. In Man a small *bile-duct* introduces the bile into the duodenum. The bile contains no digestive enzymes, but, being alkaline, it neutralizes the acid of the gastric juice and thus stops the action of the pepsin. It contains *bile-salts*, which (like soap) lower the *surface tension* of water. As a result, bile tends to keep the tiny droplets of melted fat and oil from running together, so that they present a much greater surface to the *lipase*, the fat-splitting enzyme of pancreatic juice. (See p. 122 for experiments illustrating the action of bile.)

Pancreatic juice, like bile, is alkaline. It contains three enzymes *amylase* (starch-splitting), *trypsin*† (protein-splitting), and *lipase* (fat-splitting). The starch-splitting enzyme acts on any starch that has not been split up by the ptyalin in the saliva. (Pancreatic amylase, however, can also act on uncooked starch, while ptyalin only acts on starch after the starch-grains have been burst by cooking.) The protein-splitting enzyme—trypsin—carries the digestion of proteins a stage further than pepsin, some amino-acids

being formed. The fat-splitting enzyme—lipase—hydrolyses oils and fats into fatty acids and glycerine.

Digestion is completed by the action of the enzymes of the *intestinal juice*, produced by cells in the walls of the small intestine. Intestinal juice contains at least four more enzymes. One (*erepsin*†) completes the digestion of proteins to amino-acids, while three others, *sucrase*,† *lactase*,† and *maltase*, split up the complex sugars sucrose, lactose, and maltose into simple sugars—mainly *glucose*.

A great deal of the food of plant-eating animals consists of cellulose. None of the higher animals produces a cellulose-digesting *enzyme*. Thus in Man and many carnivorous animals no use can be made of this carbohydrate as food, and it is passed out in the faeces. In most herbivorous animals, however, the cellulose is digested in the alimentary canal by *bacteria* that live there, and some of the sugars so formed are absorbed by the animal host.

It should be noted that water plays a large part in digestion. It is *secreted* with the enzymes and is, of course, replaced during drinking. The various enzymes dissolve in water and are thus brought to the surface of the solid food particles or of the fat droplets. Moreover, the digestive processes that we have called 'splitting' reactions are all examples of *hydrolysis*—water actually entering into the chemical reactions, thus:

$$\text{STARCH} + \text{WATER} \xrightarrow{\textit{amylase}} \text{SUGAR}$$

$$\text{PROTEINS} + \text{WATER} \xrightarrow{\textit{proteases}} \text{AMINO-ACIDS}$$

$$\text{FATS} + \text{WATER} \xrightarrow{\textit{lipase}} \text{GLYCERINE} + \text{FATTY ACIDS}$$

The products of digestion dissolve in the excess water and, after their absorption from the small intestine, the bulk of the water is itself taken back into the blood-stream by absorption from the large intestine.

THE FINAL PRODUCTS OF DIGESTION

As a result of digestion in the mouth, stomach, and small intestine: (i) *proteins* are finally split up into *amino-acids*, (ii) *carbohydrates* are finally split up into *simple sugars*, and (iii) *fats* are

split up into *fatty acids* and *glycerine*, the fatty acids combining with the *bile salts* in the intestine to form soluble compounds. All these substances are now ready for *absorption*.

ABSORPTION

Absorption of food takes place from the small intestine. In Man the walls of the small intestine are covered with large numbers

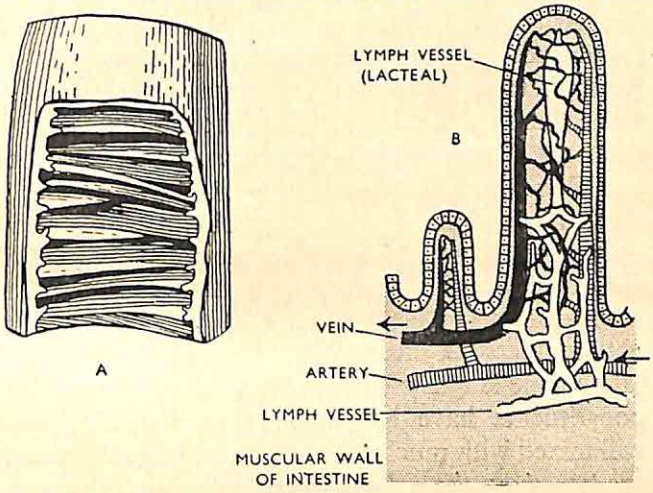


Fig. 39. Structure of human small intestine. A. Intestinal wall cut away to show transverse folds. B. Highly magnified section through villus showing blood-vessels and lymph-vessels.

of tiny, finger-shaped projections called *villi*,† each villus being about 0·05 in. long and containing a network of capillary blood-vessels running just beneath its thin covering layer of cells (see Fig. 39B). In the centre of each villus is a blind-ended tube called a *lacteal*,† which is part of the lymphatic system of the body. These villi present a very large surface area to the digested food in the small intestine, and it is through the cells covering the villi that the digested food enters the true interior of the body. The surface area of the lining of the small intestine is further increased by transverse folds (see Fig. 39A). The *simple sugars* (e.g. glucose) and the *amino-acids* pass directly into the capillary blood-vessels running through

the villi. All the veins draining blood from these capillaries in the villi finally join together to form one large vein (the *hepatic portal vein*),† which runs to the *liver* and then divides again into capillaries. (N.B.—It is unusual for a vein to divide up into capillaries *at both ends*.) Some of the digested fat is absorbed into the *blood*-vessels of the villi in the form of fatty acids combined with bile salts and then passes along the portal vein to the liver. The rest is absorbed as tiny droplets of neutral fat and passes into the *lymph*-vessels of the villi, and each of these *lacteals* acts as a tiny *lymph-heart*, pumping lymph towards the larger lymph-vessels, which finally discharge the lymph into the large vein from the left arm, at the base of the neck. The droplets of fat are then carried round the body in the blood-stream.

The excess fat is removed from the blood-stream and stored under the skin and around the organs in the body-cavity. When the animal is short of food for a long time, it makes use of this reserve fat as a source of energy.

By the time that the creamy emulsion of digested food has travelled the whole length of the small intestine, practically all the soluble constituents have been absorbed, and the indigestible residue, still mixed with much water, is passed on into the first part of the *large intestine*—the *colon*—where much water is re-absorbed from the indigestible food-residues, which are reduced to about one-tenth of their former volume. The large intestine also gets rid of some waste products from the blood. The semi-solid food-waste—or *faeces*—collects in the looped part of the large intestine, down the left side of the body-cavity, until there is a sufficient quantity to push some of it into the *rectum*, the last portion of the food-canal. In a normal, healthy person, the presence of faeces in the rectum should at once produce a desire to evacuate* the faeces through the anus.

DIFFUSION AND DIALYSIS

In order to understand how the final products of digestion pass through the walls of the villi, it is necessary to know more about

diffusion and similar processes. If a jar of oxygen is placed mouth-to-mouth with a jar of nitrogen, the two gases soon form a uniform mixture by *diffusion*, the gas molecules moving about in all directions until the two gases are thoroughly mixed. If a test-tube is half-filled with copper sulphate solution and is then *slowly* filled up with water, *without shaking*, the blue colour *gradually* spreads throughout the tube until the colour is uniform. The copper sulphate molecules have *diffused* through the water in the tube. Diffusion in liquids, however, is much slower than diffusion in gases. If a solution of egg-white is used instead of copper sulphate in the last experiment, diffusion is still slower, because the protein molecules of the egg-white are very much bigger than the copper sulphate molecules.

If a solution of sugar in water is enclosed in a membrane (e.g. parchment-paper or Cellophane) surrounded by water, some sugar molecules diffuse through the membrane into the water. This process is called *dialysis*.† At the same time, some water molecules diffuse in the opposite direction through the membrane into the sugar solution. This process is called *osmosis*. When sugar solution is put inside the bulb of a thistle funnel with its wide end closed by a membrane and standing in water, water diffuses inwards across the membrane (by osmosis) and sugar diffuses outwards (by dialysis) (see Fig. 40). The explanation of this is that the moving molecules of water and sugar are constantly hitting against the membrane, which behaves like a sieve with very tiny holes. The water molecules, being much smaller and moving more rapidly than those of sugar, find their way through the sieve much more easily, so that more water passes inwards through the membrane

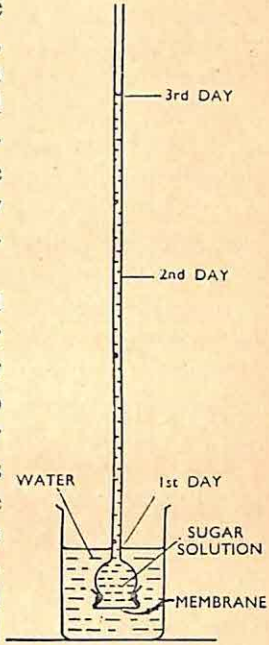


FIG. 40. Osmosis.

into the sugar solution than sugar molecules pass outwards through the membrane into the water, i.e. in this case, osmosis is more rapid than dialysis. Hence the level of the liquid inside the thistle funnel rises.

Thomas Graham (1805–69), while experimenting with this diffusion of dissolved substances through membranes, found that some substances, e.g. salt, sugar, and urea, *dialysed* (or passed through membranes), while others, e.g. egg-white, starch, gelatine, and glue, did not dialyse. As the first class of substances crystallize readily, Graham called them *crystalloids*.† The second class of substances, whose molecules are too big to pass through membranes, he called *colloids*. Animal and plant material consists of both crystalloids and colloids. For example, the proteins and starches are *colloidal* and so cannot pass through the walls of the small intestine until their large colloid molecules have been broken down into simpler, smaller, diffusible molecules of *crystalloids* (the amino-acids and simple sugars). (See pp. 125–6 for experiments on dialysis.)

Digestion, therefore, prepares the food for absorption by converting the large molecules of colloids into smaller molecules of crystalloids that can diffuse through the walls of the intestine.

ASSIMILATION AND THE FUNCTIONS OF THE LIVER

After absorption, the soluble food materials are carried by the blood-stream to the body-cells where respiration goes on and where the foods become part of the protoplasm or are '*assimilated*'. The food passes first, however, into the liver, all the veins from the food-canal joining together to form one large vein (the hepatic portal vein), which runs to the liver and there divides up into capillaries again. The liver is the largest gland in the body; it is situated in the body-cavity near the diaphragm and partly overlaps the stomach. Its most important function seems to be regulation of the composition of the blood, in fact, it is the main storehouse and laboratory of the body: for example:

1. All the sugar formed by digestion of carbohydrates is carried

by the portal vein from the small intestine to the liver. This sugar is usually in excess of immediate requirements and most of it does not pass straight through the liver into the blood-stream to the tissues, but is *stored as glycogen* (or 'animal starch'). When the main blood-stream becomes less rich in sugar (e.g. after exercise), some of the glycogen in the liver is reconverted to sugar, which passes into the blood. In this, and in other ways, the liver controls the amount of sugar in the blood.

2. The liver does not store sufficient sugar to keep a starving man's blood-sugar at normal level for long. But it can produce sugar in another way—by breaking down amino-acids brought to it by the portal vein. These amino-acids, when in excess of immediate requirements, are *not stored* in the liver, but are broken down to liberate sugar and ammonia. The ammonia thus formed is converted in the liver by a complex process (involving its combination with carbon dioxide) to *urea*. This harmless, neutral substance is passed into the blood-stream and is later *excreted* by the kidneys in the *urine*. Of course, not all the amino-acids are broken down in this way—some pass through the liver and are used to build up proteins in other body-tissues. But it is important to realize that protein as well as carbohydrate foods can be used by the body as a source of the sugar used in respiration.

3. The liver produces *bile*, whose part in digestion has already been briefly discussed on p. 113.

4. Metabolism in the liver is very active and this organ is, in fact, one of the major sources of heat in 'warm-blooded' animals.

PRACTICAL WORK ON DIGESTION

EXPERIMENTS ILLUSTRATING DIGESTION IN THE MOUTH

(i) Place small pieces of red and blue litmus-paper (or neutral, purple litmus-paper) on the tip of the tongue. Notice that, in most individuals, saliva is very slightly alkaline.

(ii) Prepare a solution of saliva as follows: Warm 50 c.cm. of water in a beaker to about 37° C. and then wash out the mouth thoroughly with about 20 c.cm. of this warm water. Then take

another 20 c.cm. of warm water in the mouth and keep it moving with the tongue for at least one minute. Collect this saliva solution in a clean beaker and filter it. Place test-tubes of saliva solution and 1 per cent. neutral starch solution (separately) in a water-bath at about 37° C. to warm up to body-temperature for your experiments. On a white tile (or a glass sheet over a piece of white paper) place drops of iodine solution (N/20) at regular intervals as on p. 80.

(iii) Put about 3 c.cm. (1 in. in a $\frac{5}{8}$-in. test-tube) of the warm 1 per cent. starch solution in a test-tube and add an equal volume of warm saliva solution. Note the time. Shake up and place in the 37° C. water-bath.

After 30 seconds remove a drop of the mixture with a glass rod and add it to one of the drops of iodine solution on the white tile. A blue colour shows that unchanged starch is still present, a purple or reddish colour shows the presence of dextrin, while a negative result shows that the starch is completely hydrolysed. Remove drops of the mixture every 30 seconds until there is no colour change with iodine.[1] If this point is reached in less than 2 minutes, dilute some of your saliva solution with an equal volume of water and repeat the experiment.

(iv) When you have reached the colourless point with iodine add an equal volume of diluted Fehling's solution (1/5) and heat in a boiling water-bath. A red precipitate of cuprous oxide shows the presence of a reducing sugar. This is because the enzyme *ptyalin* in saliva only begins the digestion of starch, converting it to maltose, which is later hydrolysed to glucose in the small intestine.

(v) Repeat these experiments, using diluted saliva solution that has been boiled *before* the starch solution is added. What conclusion can you draw from your results?

EXPERIMENTS ILLUSTRATING DIGESTION IN THE STOMACH

(i) *To show the digestion of Protein by Pepsin.* Put 4 c.cm. of

[1] Wash the glass rod under the tap between successive tests.

1 per cent. pepsin[1] solution in a test-tube and add 1 c.cm. of 1 per cent. hydrochloric acid. (1 c.cm. of 1 per cent. hydrochloric acid has now been diluted to 5 c.cm., so that the mixture contains 0·2 per cent. of hydrochloric acid, i.e. the same percentage as in human gastric juice.) Add a *small* piece of washed white fish-muscle and place the test-tube in the 37° C. water-bath. (If a *small* piece of fish is used, it will be completely dissolved and there can be no doubt about the action of the enzyme on the protein. If a large piece of fish is used, it will not *all* dissolve, and it will be difficult to be sure from direct inspection whether *any* has dissolved.) Shake occasionally (to imitate the movements of the stomach) and notice that the protein gradually dissolves, i.e. it is *digested* by the pepsin. Filter the final solution and divide it into two. To one half add 2 c.cm. of 40 per cent. sodium hydroxide solution containing 2 drops of 1 per cent. copper sulphate solution. The characteristic purple Biuret reaction (see pp. 94–95) shows the presence of dissolved proteins. To the other half add about 1 c.cm. of 1 per cent. *bile-salt* (or ox-gall) solution. A white precipitate of partly-digested protein is formed. (Bile-salts precipitate *proteoses* in the small intestine, thus giving the pancreatic juice a better chance to complete the digestion of proteins in the food.)

(ii) *To show the action of Rennin on Milk.* Take about 5 c.cm. of fresh cow's milk in a test-tube and add about 2 c.cm. of rennin solution (e.g. 'essence of rennet'). Shake, and keep the tube in the 37° C. water-bath for 5 minutes. Notice that a *clot* is formed, and that the tube can be turned upside-down without any milk flowing out. The clot consists of the *casein*† of the milk together with the *milk-fat*, which is carried down with the casein. Put back the tube in the 37° C. water-bath for an hour and notice how the clot gradually contracts, squeezing out a clear liquid, '*whey*'.† (This is the first stage in the process by which milk is made into cheese.) The same clotting takes place in the stomach when milk is drunk. If the milk is swallowed quickly, it forms one large, tough, clotted

[1] Pepsin, like some other enzymes, can be extracted from the stomachs of animals killed for food and preserved as a dry solid.

mass, which is digested only very slowly. Hence it is better to drink milk slowly, sipping a little at a time (or, better still, through a clean drinking-straw) so that loose, easily digested *curds*† are formed in the stomach.

(iii) *To show that Rennin is destroyed by boiling.* Repeat the last experiment, using 1 c.cm. of *boiled* rennin solution. No clot is formed, showing that rennin (like any other enzyme) is destroyed by boiling.

EXPERIMENTS WITH BILE

To show how Bile lowers the surface tension of Water. The following experiments may be done with diluted ox-bile[1] (1 volume of bile to 9 volumes of water), or with a 1 per cent. solution of bile-salts.

(*a*) Take two 100-c.cm. glass beakers. In one put 10 c.cm. of distilled water and in the other put 10 c.cm. of bile solution. Sprinkle a little 'flowers of sulphur' on to the liquid surface in each beaker and then rock the vessels gently. The very fine particles of sulphur continue to float on the surface of the pure water owing to its high surface tension, but the sulphur sinks in the bile solution because the bile-salts have lowered the surface tension of the water.

(*b*) Put two similar folded filter-papers in two filter-funnels. Wet one with pure water and the other with bile solution. Then pour about 1 c.cm. of vegetable oil, e.g. coconut oil, on to each filter and allow to stand for some hours. Notice that the oil will not pass through the filter-paper moistened with pure water, owing to the high surface tension of the oil–water surface. The oil, however, gradually passes through the filter-paper wetted with bile solution, showing that the bile-salts reduce surface tension. This fact is of great importance in breaking up oils into very tiny drops so as to form an *emulsion*, and this is an important function of bile in the digestion and absorption of fats.

EXPERIMENTS WITH EMULSIONS

Take four test-tubes, A, B, C, and D. Put 10 drops of vegetable

[1] Obtained from the abattoir.

oil, e.g. coconut oil, in each tube. To A add 10 c.cm. of distilled water. To B add 5 c.cm. of distilled water and 5 c.cm. of sodium carbonate solution (1 per cent.), giving about the same amount of alkali as in human pancreatic juice. To C add 10 c.cm. of 1 per cent. bile-salt solution. To D add 5 c.cm. of 1 per cent. sodium carbonate solution and 5 c.cm. of 1 per cent. bile-salt solution. Shake all the tubes vigorously and allow them to stand in the test-tube rack. In each case, shaking breaks up the oil into very fine drops that are suspended in the liquid to form an *emulsion*. On standing, however, the oil-drops in A soon run together and rise to the surface to form a separate layer, i.e. the emulsion is only *temporary*. In B the emulsion is *permanent*, because the alkali forms a trace of soap that produces a protective film round the oil-drops. (The pancreatic juice is alkaline, and thus helps to break up oils and fats into small droplets.) In C a permanent emulsion is formed because the bile-salts lower the surface tension between the oil and the water. In D a very stable emulsion is formed by the combined action of the alkali and the bile-salts (both of which are present in the small intestine).

EXPERIMENTS ILLUSTRATING PANCREATIC DIGESTION

To show the action of Pancreatic Juice on Starch. *Amylase*, the starch-splitting enzyme of pancreatic juice, does not act properly except in the presence of certain salts; hence in making up the 'pancreas solution' for these experiments on the digestion of starch, add 0·8 gm. of sodium chloride and 1·2 gm. of sodium phosphate ($Na_2HPO_4.12H_2O$) to every 100 c.cm. of 'pancreas solution'. In a test-tube put 3 c.cm. of the 'pancreas solution', 3 c.cm. of 1 per cent. starch solution, and 2 c.cm. of 2 per cent. sodium carbonate solution. Shake, and place the tube in the 37° C. water-bath. Remove drops of the mixture at one-minute intervals and test for starch with drops of iodine solution on a white tile as in your experiments on salivary digestion. (N.B.—Use N/20 iodine solution acidified with dilute hydrochloric acid, since starch gives no blue colour with iodine in alkaline solutions.) After

20 minutes add a little diluted Fehling's solution and heat in a boiling water-bath. A yellow or orange precipitate shows that the starch has been converted into a reducing sugar.

To show the action of Pancreatic Juice on Protein. Mix together 3 c.cm. of 1 per cent. 'pancreas solution' (without sodium chloride or phosphate) and 3 c.cm. of 1 per cent. sodium carbonate solution in a test-tube, and place in the 37° C. water-bath. Add a *small* piece of washed white fish-muscle and shake from time to time (to imitate the movements of the small intestine). Notice that the protein gradually dissolves, i.e. it is digested by the trypsin of the pancreatic juice.

Filter the solution and divide it into two. To one half, add 2 c.cm. of 40 per cent. sodium hydroxide solution containing 2 drops of 1 per cent. copper sulphate solution. A *pink* colour shows that trypsin carries the digestion of proteins a stage further than pepsin. (Pepsin changes proteins into *proteoses* and *peptones*, which give a *purple* colour with the Biuret test, while trypsin converts proteins to *polypeptides* and *amino-acids*, and the polypeptides give a *pink* colour with the Biuret test.) To the other half of the filtrate add a few drops of 1 per cent. bile-salt solution. There is no precipitate, showing the absence of proteoses. (In the corresponding experiment on the filtrate from pepsin digestion on p. 121 there *was* a precipitate, showing that pepsin only digests proteins as far as proteoses.)

To show the action of Lipase on Fats. Take about 15 c.cm. of fresh cows' milk (which contains fat in a very finely divided condition) in a test-tube and add a few drops of 1 per cent. phenolphthalein† solution. Then add 1 per cent. sodium carbonate solution, drop by drop, shaking after adding each drop, until a permanent pink colour is obtained, showing that the mixture is slightly alkaline. Pour 5 c.cm. of the mixture into each of three test-tubes—A, B, and C—and place the tubes in the 37° C. water-bath. Take about 5 c.cm. of 1 per cent. 'pancreas solution' and make it slightly alkaline by adding 1 per cent. sodium carbonate solution, drop by drop, until just pink. To the alkaline milk in A

add 2 c.cm. of water. To B add 2 c.cm. of the alkaline 'pancreas solution' and shake. To C add 2 c.cm. of alkaline 'pancreas solution' that has been boiled. Tube B soon loses its pink colour, showing that the lipase has begun to split up the milk-fat into glycerine and fatty acid. This free fatty acid neutralizes the sodium carbonate in the mixture, which thus becomes acid. Add more 1 per cent. sodium carbonate solution to B, drop by drop, until it again turns pink, and then replace in the water-bath. The pink colour soon disappears again as more fatty acid is set free by the action of the lipase on the milk-fat. Tube C remains pink, showing that the fat-splitting enzyme was destroyed by boiling.

Milk was used in the above experiment because it contains fat in very tiny droplets (i.e. in the form of an *emulsion*) that are readily attacked by the enzyme. Other fats and oils may be used, but digestion is much slower unless the fat is finely divided. Instead of milk, an emulsion of egg-yolk can be used (1 volume of egg-yolk shaken up with 9 volumes of water). To carry out a similar experiment with a vegetable oil, e.g. coconut oil, take 5 c.cm. of absolute alcohol in a dry test-tube, add 10 drops of the oil, and warm gently until a clear solution is obtained. Pour this alcoholic solution of oil into 10 c.cm. of water. The oil is thrown out of solution in very tiny droplets that form a milky emulsion.

EXPERIMENTS ON DIFFUSION

One-third fill a 'sausage-skin' parchment-paper or Cellophane sac with 1 per cent. starch solution and then fill up the remaining two-thirds with 5 per cent. glucose solution. Stand the sac in a narrow jar or tall beaker (see Fig. 41). After 20 minutes pour a few c.cm. of water from the beaker into a test-tube and add a drop of iodine solution. No blue colour develops, showing that no starch (a *colloid*) has *dialysed* outwards through the membrane. Pour out another 2–3 c.cm. of water from the beaker and test for

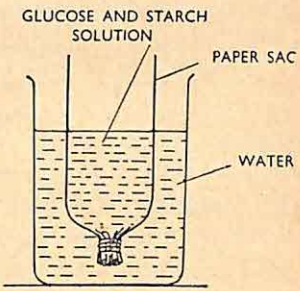


FIG. 41. Dialysis.

glucose with Fehling's solution. A red precipitate of cuprous oxide shows that some glucose (a *crystalloid*) has dialysed outwards through the membrane. Add a few drops of iodine solution to the remainder of the water in the beaker and set aside for half an hour. A blue colour develops *inside* the membrane, showing that

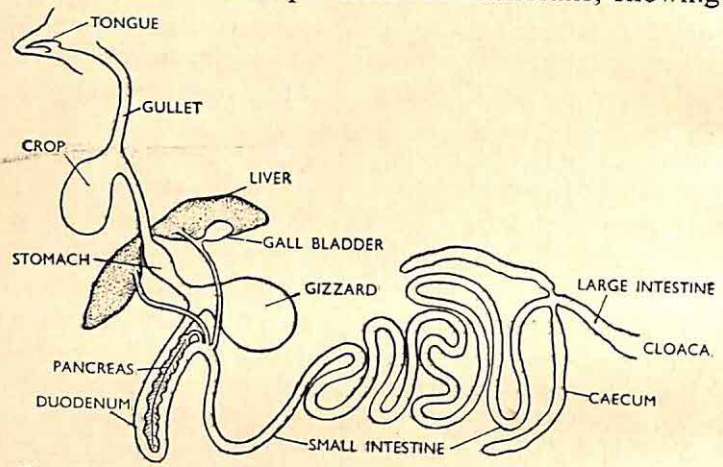


FIG. 42. Food-canal of domestic hen (diagrammatic—after Pincher).

iodine (a crystalloid) has dialysed through the membrane. No blue colour develops in the surrounding water, showing that no colloidal starch has dialysed outwards through the membrane.

Digestion, therefore, prepares the food for absorption by (i) *converting insoluble materials into soluble substances* (e.g. starch to sugar), (ii) *converting the large molecules of colloids into smaller molecules of diffusible crystalloids* (e.g. proteins to amino-acids), (iii) *converting soluble substances into substances acceptable to the tissues* (e.g. maltose to glucose).

NUTRITION IN OTHER ANIMALS

NUTRITION IN BIRDS (e.g. PIGEON)

Since a bird has no teeth, it cannot chew its food, but swallows it immediately. Hence a bird's salivary glands are of little impor-

tance for digesting carbohydrates, although the saliva lubricates the food and makes it easy to swallow. The food is passed down the gullet, the posterior part of which is enlarged to form a large, thin-walled *crop* in which food is stored temporarily. The food then passes, a little at a time, into the *stomach*, where it receives a digestive juice, and then passes (in most birds) into the *gizzard*, which has very strong, muscular walls and contains small stones that help to grind up the food (see Fig. 42). The small intestine, which is long and much looped, is connected by ducts to the liver and pancreas, receiving bile and pancreatic juice from these glands. The large intestine is very short and ends in a *cloaca*, which also receives the urinary and reproductive ducts. The details of digestion, absorption, and assimilation are much the same in both birds and mammals. Birds have very efficient digestive systems, and the amount of waste matter passed out as faeces is only small.

NUTRITION IN AMPHIBIANS (e.g. FROG)

The frog has a long tongue attached to the *front* of the mouth-cavity. The animal shoots this tongue out of its mouth suddenly, catching small animals (e.g. insects, worms, and snails) with the sticky end and then drawing the tongue back into its mouth. The food is usually swallowed whole, and there are no salivary glands. After being forced down the gullet into the stomach, the food is acted on by the gastric juice, forming a soft pulp that passes through a ring of muscle into the small intestine. The bile-duct, leading from the gall-bladder, opens into the beginning of the small intestine. The pancreatic ducts also join the bile-duct, so that both bile and pancreatic juice enter the small intestine through the same duct (see Fig. 43). The small intestine is a long, narrow tube, much looped, thus providing a large area for absorption. It widens to form the large intestine, which is comparatively short, and the faeces are passed out through the cloaca, which also receives the urinary and reproductive ducts. The chemical processes involved in digestion, absorption, and assimilation in the frog are very similar to those in mammals and birds.

NUTRITION IN FISHES

Digestion in fishes is very similar to digestion in the frog. Fishes are provided with teeth and movable jaws. The food is taken into the mouth and passed through a ring of muscle into the gullet, comb-like *gill-rakers* preventing food from passing out through the

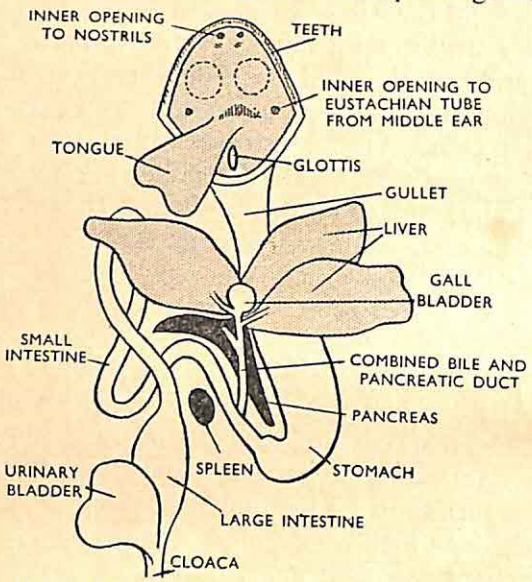


FIG. 43. Food-canal of frog (diagrammatic—after Hatfield).

gill-slits on the way. The general arrangement of the digestive organs and the processes of digestion and absorption show no fundamental differences from those of the frog (see also Book Three, Chap. III).

NUTRITION IN MOLLUSCS (e.g. SNAIL)

The snail has a horny jaw-plate in the roof of its mouth and a file-like 'tongue' in the floor of its mouth. The 'tongue' is covered with thousands of tiny, backward-pointing 'teeth', and this flexible 'file' is drawn backwards and forwards over the plants on which the animal feeds. The finely-divided food is mixed with saliva in

the mouth and is passed down the gullet into the large, thin-walled crop, and from thence into the 'stomach' and intestine (which are embedded in the 'liver'). The undigested food-remains are then passed through the large intestine to the anus (see Fig. 113).

NUTRITION IN INSECTS (e.g. COCKROACH)

Insects show great variety in their feeding habits, and their mouth parts and digestive organs are modified to suit the insect's

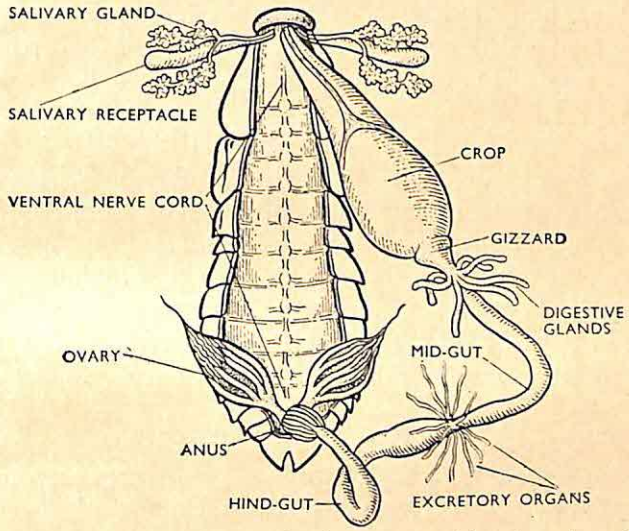


FIG. 44. Digestive and reproductive systems of female cockroach (after Bourne).

food. We shall consider only one example—the cockroach. The cockroach crushes its food with its three pairs of mouth parts, mixes it with saliva, and forces it into its mouth and down a short gullet into the large, thin-walled '*crop*', where the food is mixed with digestive juices so that digestion of carbohydrates, fats, and proteins begins in the 'crop'. Behind the 'crop' is the '*gizzard*', a round, muscular organ in which the food, mixed with digestive juices, is ground up by means of a ring of six horny 'teeth'. The 'gizzard' leads to the 'stomach' (or *mid-gut*), where the food is mixed with a

digestive juice supplied by seven or eight small finger-like digestive glands, which are extensions of the food-canal (see Fig. 44). Digestion and absorption are completed in the mid-gut and its side tubes (the only part of the food-canal that is not lined with horny material like that on the outer surface of the animal) and the undigested food-remains pass through the *hind-gut* to the anus. The digestive enzymes of the cockroach enable it to digest almost any animal matter or vegetable matter, e.g. paper, cotton, wool, silk, leather, hair, and horn. The cockroach usually stores large reserves of fat in the body-cavity. The absorbed food reaches the tissues by way of the blood that fills the body-cavity.

NUTRITION IN SEGMENTED WORMS (e.g. EARTH-WORM)

The food-canal of the earth-worm consists of a straight tube, running through the middle of the body-cavity from the mouth to the anus: the simplest possible food-*canal* (see Fig. 112). Earth-worms eat their way through the soil, sucking in and swallowing soil containing particles of animal and vegetable matter. The soil is ground up by the '*gizzard*' and is then passed through the intestine, where digestion and absorption take place.

NUTRITION IN TWO-LAYERED ANIMALS (e.g. HYDRA)

Hydra has no food-*canal* but only a food-*cavity* surrounded by a double layer of cells (see Fig. 111). There is no *body*-cavity between the food-cavity and the body-wall as there is in all the higher animals. Living food-organisms are paralysed by the stinging-cells and are swept into the 'mouth' by the hollow threads (or *tentacles*†) that surround the upper, open end of the animal. Some of the cells of the inner layer of the food-cavity are *gland-cells* that pass out digestive juices into the food-cavity, where digestion of the larger food-particles begins. After these larger particles have been broken up by enzymes, the smaller particles are *ingested* by Amoeba-like cells, and digestion is completed by these cells. Undigested food-remains are got rid of through the only opening to the food-cavity—the 'mouth'. Hydra is an excellent example of

the beginnings of '*division of labour*' between different kinds of specialized cells. For example, there are several different kinds of cells taking part in the one function of nutrition: gland-cells produce enzymes, Amoeba-like cells ingest food-particles and complete the process of digestion, stinging-cells paralyse the prey, muscle-cells move the tentacles, and nerve-cells control the other kinds of cells.

NUTRITION IN ONE-CELLED ANIMALS (e.g. PARAMECIUM AND AMOEBA)

Paramecium shows the beginnings of a definite path for the food through the body during digestion. There is a groove on one side, leading to a funnel-shaped opening into the softer protoplasm underlying the firmer outer layer (see Fig. 45). Food-particles, e.g.

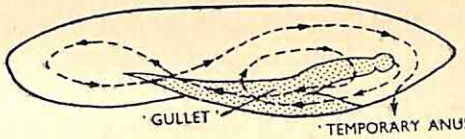


FIG. 45. Food-path in Paramecium (diagrammatic—after Borradaile).

bacteria, are swept into this groove (or 'gullet'), and enter the protoplasm enclosed in a drop of water. Owing to the streaming movement of the protoplasm, these food-particles circulate round the body, following a definite path, while digestion and absorption take place. The undigested residue is finally got rid of, always at the same point in the surface. Although no permanent opening is visible at this spot, it can be regarded as a 'temporary anus'.

Nutrition in Amoeba is even simpler than in Paramecium. Food can be ingested *at any point* in the Amoeba's surface, there is *no definite path* along which the food moves during digestion and absorption, and the undigested food-residues are got rid of *at any point* in the surface, i.e. there is no one spot serving as a 'temporary anus' (see Fig. 46). In spite of the extreme simplicity of its method of nutrition Amoeba carries out the same fundamental processes as the higher animals. After *ingestion*, the food-particle is acted

on by an acid digestive juice followed by an alkaline digestive juice so that *digestion* of the food takes place. *Absorption* and *assimilation* by the protoplasm complete the process of nutrition.

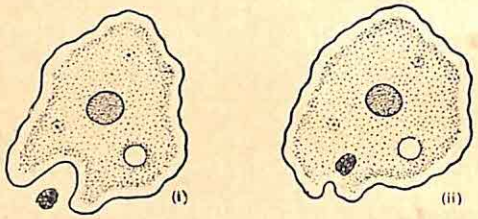


FIG. 46. An Amoeba ingesting a food particle (highly magnified).

SUMMARY OF NUTRITION IN ANIMALS

It should now be clear that, in all animals, *nutrition* consists of the same fundamental processes of *ingestion*, *digestion*, *absorption*, and *assimilation*. All animals feed on 'ready-made' complex organic food in many different forms, but the carbohydrates, fats, and proteins of the food are split up into simpler, soluble, and diffusible substances that can be absorbed and assimilated by the body and used for growth and repair or for supplying energy.

(See *General Science Workbook*, Book Two, pp. 76–81, for learning exercises on this chapter.)

Chapter VI

RESPIRATION IN PLANTS AND ANIMALS

One of the things that food does is to provide the living organism with *energy*. A steam engine gets its energy by burning (or oxidizing) fuel. A living thing gets its energy in a similar way—by oxidizing part of its food. This process, which is called *respiration*, goes on at normal temperatures. Some of the energy in the food molecules is set free as heat, some as mechanical energy, and some is used in driving anabolic chemical reactions, i.e. it is locked up again in new and more complex chemical molecules.

TISSUE RESPIRATION

The essential process of respiration—sugar oxidation with transformation of energy—goes on in all the living, individual cells of the plant or animal. It is called *tissue-respiration*† or *internal-respiration* (to distinguish it from *breathing* or *external respiration*). As our earlier experiments have shown (Book Two, Chap. V), one end-product of respiration is carbon dioxide, and oxygen is used up. Sugar is oxidized by a complicated series of chemical reactions that are probably closely similar in all living cells. During recent years scientists have discovered a great deal about these reactions, some of which can go on outside the cells as long as the proper *enzymes* are added, but it is not possible in this Course to give even an elementary account of the *chemistry* of respiration. We can say something, however, about the *respiratory mechanisms*—the ways in which the cells of different organisms obtain the oxygen that they use and give out the carbon dioxide that they produce during respiration.

RESPIRATORY MECHANISMS IN ANIMALS

In studying nutrition we started with the 'highest' plants and animals and worked downwards to the 'lowest'. We shall begin

our study of respiratory mechanisms in the 'lowest', one-celled animals and work upwards to Man.

RESPIRATION IN AMOEBA AND PARAMECIUM

In a tiny, one-celled animal like Amoeba or Paramecium there is no special respiratory system. The animal is surrounded by water containing dissolved oxygen that can be absorbed *over the whole surface*, and every part of the protoplasm can get the oxygen it requires *by diffusion.*

RESPIRATION IN HYDRA

Although Hydra shows the beginnings of 'division of labour' between different kinds of specialized cells, there are *no special respiratory organs.* As Hydra is never more than two cells thick, oxygen can be absorbed from the surrounding water over the whole surface and can reach every part of the body *by diffusion*, as in the simplest one-celled organisms.

RESPIRATION IN THE EARTH-WORM

The earth-worm has *no special breathing organs* and absorbs oxygen *over its whole body-surface*, through its *damp skin.* But an earth-worm is too 'thick' for oxygen to reach every cell quickly enough by diffusion; hence a *transport system* is necessary. The earth-worm has a simple blood-system for this purpose. Underneath the skin is a network of capillary vessels, containing blood. This blood is coloured red by its *dissolved haemoglobin* (a protein substance containing iron), which is the 'oxygen-carrier' of the blood. The oxygen absorbed through the skin combines with this haemoglobin, forming *oxy-haemoglobin.* This loosely-combined oxygen has to be distributed to every part of the worm, and the carbon dioxide produced during respiration has to be got rid of through the skin; hence the blood *circulates.* In the earth-worm there is a closed system of blood-vessels, some of which contract and expand regularly (like hearts) and thus cause the blood to circulate round the body (see Fig. 112). Part of the blood filters through the walls of the capillaries and bathes every cell as *lymph.*

RESPIRATION IN INSECTS

In the insect body (see Fig. 47) a complex branching system of fine tubes, opening to the air through breathing-pores, carries air to every part of the body: and the final exchange of gases with the cells takes place by diffusion. Blood is not concerned with oxygen transfer in insects. The gases taking part in respiration (oxygen and carbon dioxide) enter and leave the finer branches of the air-tubes by diffusion, but there is also, very often, a true *breathing* action. The insect carries out regular muscular movements and these, together with the normal body movements, cause expansion and contraction of the air-tubes, so that air is actually 'sucked' into and pushed out of the larger air-tubes. It is interesting to note that just as plants lose *water vapour* as well as carbon dioxide through their stomates, so insects lose water vapour through the breathing-pores of their air-tubes.

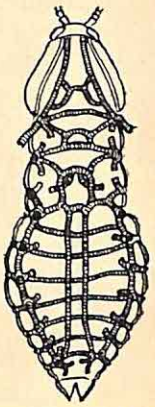


FIG. 47. Main air-tubes of cockroach (after Bourne).

Most insects live on land or in the air. Some, however, live in water, and these show many curious *adaptations* in their respiratory mechanisms. In many water-insects the air-tubes mostly open at one end of the body, and this can be pushed above the surface of the water when the insect comes up for air. You will see this taking place if you watch a mosquito larva, or 'wriggler'* (see Fig. 95, Book Three). In some aquatic insect larvae the air-tubes do not open at pores but branch into still finer tubes just under the skin of special projecting flaps (gills). These insects, like the fishes, take up oxygen from the water in which it is dissolved; and are thus able to live permanently under water.

RESPIRATION IN THE SNAIL

The snail changes the air in its *mantle-cavity* from time to time (see Fig. 113). The mantle-wall is richly supplied with a network of capillary blood-vessels, so that exchange of oxygen and carbon dioxide takes place between the blood and the air (as in the *lung*

of the higher animals). The snail's blood, like that of some of the other lower animals, does not contain haemoglobin (an iron-protein substance) but *haemocyanin*† (a copper-protein substance). Haemocyanin is colourless, but after combining with oxygen the *oxy-haemocyanin* is slightly blue. The snail's heart is very simple, consisting of *two* chambers only, a thin-walled auricle and a thick-walled, muscular ventricle (see Fig. 113). The oxygenated blood from the capillaries of the mantle is collected into a large vein

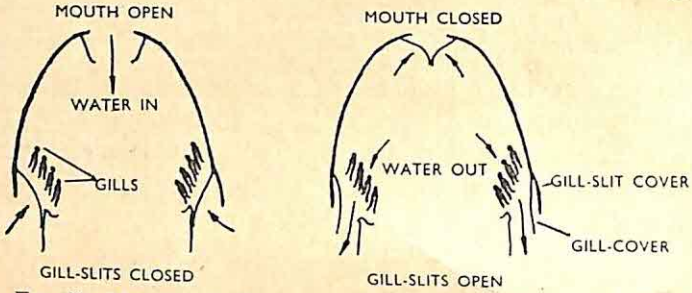


FIG. 48. Movements of mouth and gill-covers of bony fish during respiration (diagrammatic).

leading to the auricle. The ventricle then pumps this oxygenated blood along the main artery to all parts of the body.

RESPIRATION IN FISHES

Watch a goldfish in an aquarium. It seems to be constantly drinking the water. Actually it is taking water into its mouth and then expelling* it through the *gill-slits* that open at each side of the back of the mouth cavity under the *gill covers*, which you will see moving. This respiratory action is illustrated in Fig. 48. The water (containing dissolved oxygen) passes over the *gills*, which are numerous folds of very thin tissue richly supplied with capillary blood-vessels (see Fig. 49). Exchange of oxygen and carbon dioxide takes place between the water and the blood in the gill-capillaries, the oxygen combining with haemoglobin in the red blood-cells. This oxygenated blood from the gills is distributed by arteries to every part of the body and it becomes de-oxygenated in

the body-capillaries, the oxygen being used in the respiration of the body-cells. The de-oxygenated blood is then collected up by veins and returned to the heart, from which it is pumped to the

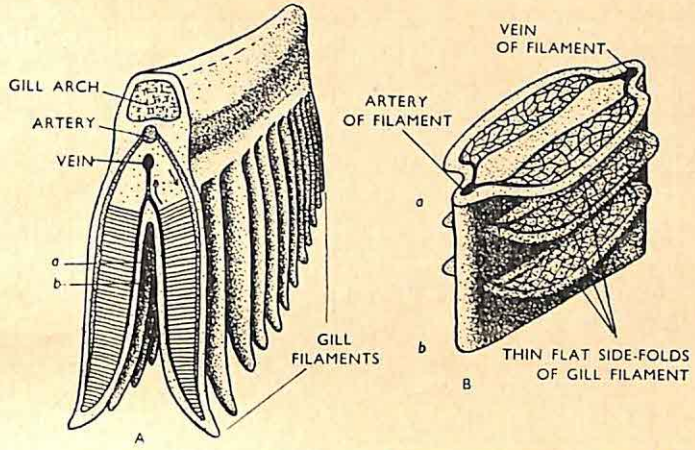


FIG. 49. Gill structure of bony fish. B is a more highly magnified view of the portion *ab* from A (after Dakin).

gills once more to complete the circulation. The heart of a fish is very simple, consisting of *two* chambers only, an auricle and a ventricle, with valves between the two to stop 'back-flow' (see Fig. 50). A fish's heart contains de-oxygenated blood only.

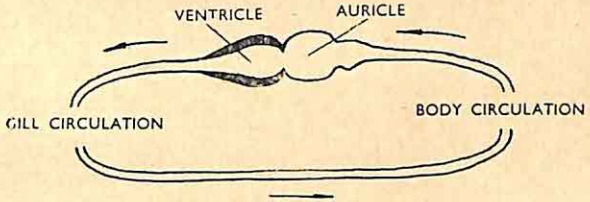


FIG. 50. Fish blood circulation (very diagrammatic).

Fishes, therefore, use oxygen that is dissolved in the water surrounding them. Strange to say, they die from *lack* of oxygen when kept out of water. This is probably because the unsupported

folds of the delicate gills fall together and thus greatly reduce the absorbing area through which oxygen can pass.

RESPIRATION IN THE FROG

The frog is of special interest since the adult animal has no less than three ways of absorbing oxygen: (*a*) *through its lungs*, (*b*) *through its damp skin*, and (*c*) *through the lining of its mouth*. Throughout its entire life-history, the frog absorbs oxygen through its skin, which has a network of capillary blood-vessels. In its earliest *tadpole* stage, however, immediately after hatching, it has three pairs of *external gills* on the sides of the body, just posterior to the head. A little later, *gill-slits* appear and the external gills wither away, being replaced by a new set of *internal gills* like those of a fish. The tadpole's blood-system, also, is like that of a fish, consisting of a simple *two*-chambered heart circulating blood through the gill-capillaries and round the body. Up to this stage in its development, therefore, the tadpole is practically a *fish*. After this stage, the developing frog gradually loses its fish-like habits and structure. Fore and hind limbs appear and the tail is absorbed. The gills wither and the gill-slits close up, being replaced by *lungs*. The blood-system also changes to suit this new method of respiration, the auricle dividing into two, producing a *three*-chambered heart. The frog's lungs are simple, hollow sacs, the thin walls of which are folded inwards, thus increasing the absorbing area. The walls of the lungs are richly supplied with blood-capillaries, and gases are exchanged between the blood and the air in the lungs. We have already discussed the way in which a frog breathes (Book Three, Chap. III). In the frog, *lymph* plays an important part in respiration, carrying oxygen from the red blood-cells through the walls of the capillaries to every cell. After exchanging oxygen for carbon dioxide and other waste products of the cells, the lymph is collected up by thin-walled lymph-vessels. There are also large *lymph-sacs* between the skin and the muscular body-wall, and these are connected with the lymph-vessels. Two pairs of *lymph-hearts* return this lymph to the veins.

The heart of an adult frog is somewhat conical in shape and it is divided into *three* chambers—a right auricle, a left auricle, and a single thick-walled ventricle. *De-oxygenated blood* from the *body-circulation* enters the *right auricle*, while *oxygenated blood* from the *lung-circulation* enters the *left auricle*. As the two auricles contract, both kinds of blood are forced into the ventricle, but there is very little mixing of the two different kinds of blood owing to the spongy structure of the ventricle. Hence, when the ventricle is full, its left side contains oxygenated blood, its right side contains de-oxygenated blood, and there is a little mixed blood in the middle. When the ventricle contracts, the blood is forced along the main artery, which branches and distributes blood to the lungs and to all other parts of the body. The three branch-arteries on each side, however, are arranged in such a way that (*a*) *the de-oxygenated blood goes to the lungs and skin*, (*b*) *the mixed blood goes to the hinder part of the body*, and (*c*) *the oxygenated blood goes to the head and brain.*

RESPIRATION IN BIRDS (e.g. PIGEON)

Birds are very active animals and have a high body-temperature, several degrees warmer than that of mammals. This uses a great deal of energy that can only be supplied by a very efficient method of breathing. The bird's lungs, which are comparatively small, are attached to the ribs; hence they do not expand and contract during breathing. But the bird's lungs are not blind sacs, they open into a system of large, thin-walled *air-sacs* lying in the body-cavity. Every time that air is breathed in, it rushes *through the lungs* into the air-sacs, and the same half-used air is later forced out from the air-sacs through the lungs again (see Fig. 51). In this way the blood in the lung-capillaries has two opportunities for absorbing oxygen from the same air during each breathing movement. During flight, the action of the flying-muscles helps the breathing movements. This very efficient breathing system helps to provide a bird with its high body-temperature and enables it to be one of the most active of all animals.

The bird's blood-system, also, is very efficient. The red blood-cells of birds are larger than those of mammals. As in mammals, there is a *four*-chambered heart, and a *double circulation* (*heart→ lungs → heart → body → heart*) which keeps the oxygenated blood

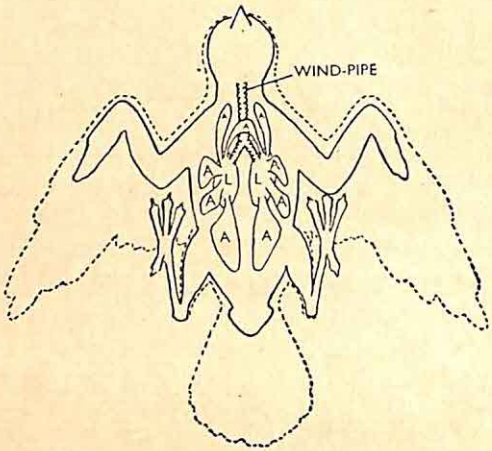


FIG. 51. Lungs and air-sacs of bird. L. lung. A. air-sac (after Cooke, Burkitt and Barker).

quite separate from the de-oxygenated blood. The right side of a bird's heart contains only de-oxygenated blood returning from the body-circulation, while the left side contains only oxygenated blood returning from the lung-circulation.

RESPIRATION IN MAMMALS (e.g. MAN)

We have already discussed the mechanism of breathing (or *external respiration*) in Man (Book Three, Chap. II). Mammals are the only animals with a *diaphragm* forming the air-tight cavity necessary for this type of breathing. (Thus birds and frogs, which also have lungs (but no diaphragm), breathe in quite a different way.) The lungs are covered with a thin membrane and they hang freely in the chest-cavity, which is also lined by a similar membrane. A film of liquid separates the two layers of membrane so that they slide easily over each other during breathing. The

system of branching air-tubes ending in air-sacs inside the lungs provides an enormous area for the absorption of oxygen by the blood in the lung-capillaries. We have already described the blood-system of the mammal (Book Three, Chap. II).

Oxygen diffuses into the blood-capillaries through the thin membrane lining the air-sacs, and combines with the haemoglobin of the red blood-cells. For oxygen to pass from an air-sac into a capillary it need diffuse through two cells only, because the walls of the air-sacs and of the capillary blood-vessels are each only one cell thick. This oxygenated blood finally reaches the capillary blood-vessels of the body-circulation. In these thin-walled capillaries the oxy-haemoglobin gives up its loosely-combined oxygen to the lymph, which escapes through the walls of the capillaries and bathes the individual body-cells. *The essential feature of respiration—the actual liberation of energy—takes place in the individual body-cells.* (This is called *internal respiration*, or *tissue-respiration*.) In the living cell the food-substance that is oxidized to liberate energy is always a simple sugar (usually *glucose*) and the oxidation is brought about by *enzymes* present in the cell. Besides liberating energy, this oxidation also produces carbon dioxide and water, which diffuse out of the cell into the lymph and from thence to the blood-capillaries. On reaching the lungs, the carbon dioxide and water diffuse into the air-sacs and are got rid of during breathing.

You should now realize that there is no essential difference between the mechanism of respiration in a one-celled organism like Amoeba and in the individual body-cells of higher animals.

BODY TEMPERATURE

Animals are sometimes classified as either '*cold-blooded*' or '*warm-blooded*'. The temperature of the blood of fishes, frogs, and snakes, for example, changes with the temperature of their surroundings, and as these animals often 'feel cold' to the touch, they are said to be 'cold-blooded'. This is a misleading term, since the blood of a snake lying in hot sunshine may rise to a high

temperature. '*Animals of variable temperature*' would be a better description of such creatures. In contrast, the 'warm-blooded' animals (i.e. birds and mammals) should be called '*animals of constant temperature*', for they are able to keep their blood (and therefore their bodies generally) at a remarkably constant temperature, whatever the temperature of their surroundings may be. A man's blood, for example, is always between 97° and 99° F. (usually 98·4° F. = 37° C.) when he is healthy, whether he is living near the Equator or inside the Arctic circle. The body-temperature of most birds is even higher, around 106° F. (= 41° C.).

This constant high temperature is of great advantage to birds and mammals, since it enables them to be equally active (e.g. in getting food) under different external conditions. We have seen that an animal gets its supply of energy by oxidizing sugar and that this chemical change goes on, with many others, in the body-cells. These chemical changes are more rapid at higher temperatures; thus muscular action is more rapid when the body-temperature is reasonably high. In a cold winter most of the land animals of variable temperature *hibernate*,† that is, they pass into a kind of prolonged sleep that usually lasts throughout the cold period. During this period of inactivity they are able to survive without eating or drinking for several months. They become active again when the air temperature rises sufficiently.

Most mammals and birds remain active throughout even very cold winters and they are able to do this because of their *constant body-temperature*.

REGULATION OF BODY-TEMPERATURE

If the temperature of the body is to be kept steady between 97° and 99° F., there must be a close balance between heat gain and heat loss.

The body *gains heat* from two sources: (*a*) from the heat produced during *internal respiration* and (to a smaller extent and under certain conditions described below) (*b*) from its surroundings. We shall see (on p. 149) that heat production due to

respiration is about 1 Calorie per kilogram of body-weight per hour, when resting, and that the amount of heat produced increases with increased muscular activity. If there were no heat loss from the body, the body-temperature would rise about 1° C. (= 1·8° F.) per hour under resting conditions and it would rise still more rapidly during muscular exercise. But our mechanism for temperature regulation is so efficient that the body-temperature is kept very nearly the same at all times. It is only during very severe exercise or when the surrounding air is much hotter than the body and nearly saturated with water-vapour that heat is lost too slowly to preserve the normal body-temperature in a healthy person. Most of the heat gained by the body is produced inside the body-cells during respiration and is carried to all parts of the body by the blood-stream. Since respiration goes on continuously, the body is always receiving heat from this internal source. The body can also take in heat from external sources under certain conditions, e.g. when the surrounding air-temperature is higher than the skin-temperature or when the body-surface is exposed to sunshine; but such external sources supply much less heat to the body than it gets from internal respiration.

The body *loses heat* mainly from the *skin*, though some body-heat is also lost during breathing (as will be explained later in this chapter). Three distinct physical processes are responsible for heat loss from the skin. In order of their relative importance under tropical conditions, these are (i) by *evaporation of water* from the skin, (ii) by *radiation* from the body-surface to cooler neighbouring objects, and (iii) by *conduction* and *convection* to the surrounding atmosphere when the air-temperature is lower than that of the skin. These methods of heat loss require further explanation.

(i) *Heat loss by evaporation of water*

The main factor in heat loss at high temperatures, or when the muscles are in very active use, is evaporation of water from the skin. Even under cool conditions the skin is always giving off a

little water-vapour, but under hot conditions, or when hard muscular work is being done, large quantities of water are lost as *sweat*, sometimes as much as a litre in an hour. This sweat escapes from the *sweat-pores* in the skin. It is formed in knots of coiled tubes—*sweat-glands*—deep down in the skin, and it passes along the *sweat-ducts* to the *sweat-pores* (see Fig. 92).

Sweat is mainly water, and the sweat-glands remove this water from the blood as it flows through the skin. The evaporation of 1 gm. of water from the skin at 97° F. requires about 0·58 Calories (= 580 'small calories') of heat. Most of this *latent heat* is taken from the skin itself, thus cooling it. Hence the more sweat is *evaporated from the skin* the more heat is lost by the body. Since evaporation is more rapid in moving air than in still air, a current of air 'feels cooler' than still air of the same temperature and humidity. It is important to realize, however, that sweating can have no cooling effect unless the sweat evaporates. When drops of liquid sweat remain on the skin without evaporating, there is great discomfort owing to the lack of any cooling effect. This is why a hot, damp atmosphere, almost saturated with water-vapour, is so much less comfortable than a dry atmosphere *at exactly the same temperature*.

Heat is also lost by evaporation of water from the moist internal surfaces of the lungs and air-passages during breathing, but the amount of heat lost in this way is not affected by changes in the external temperature until *panting** begins, exposing a larger surface area of the tongue and mouth cavity and changing the air more frequently by rapid, shallow breathing. Panting provides an important method of heat loss among mammals with thick fur and few sweat-glands, e.g. dogs. A dog pants very rapidly under hot external conditions or after severe exercise and thus increases heat loss by evaporation of water from the tongue and from the surface of the mouth cavity.

(ii) *Heat loss by radiation*

In Book Three we learnt that *radiation* is the transfer of heat

from one object to another *without any physical contact between them*. Whenever two neighbouring objects are at different temperatures, heat is *radiated* from the surface of the hotter object and is *absorbed* through the surface of the cooler object, and the greater the difference in temperature between them the greater is the amount of heat exchanged. The human body radiates heat to every cooler object in its neighbourhood, e.g. to the furniture, walls, floor, and ceiling of a room, and this is the main factor in heat loss from the body *at low temperatures*. When the air-temperature is between 70° and 80° F. (= 21° to 27° C.) about two-thirds of the total heat loss takes place by radiation, but once the air-temperature exceeds the skin-temperature then there can be no cooling by radiation, only by evaporation of water from the skin.

(iii) *Heat loss by conduction and convection*

Conduction is the transfer of heat to any cooler substance *in contact* with the body, e.g. air that covers the skin and air that enters the lungs and air-passages during breathing receives heat from the body *by conduction* if the surrounding atmosphere is below body-temperature. Conduction is a much less important means of heat loss than either evaporation or radiation. The skin of most mammals is covered by a layer of hair or fur and that of birds by feathers. These body-coverings entrap air to form a heat-insulating layer over the skin. Man has little body-hair and in cold climates he has to prevent dangerous heat loss by wearing insulating clothes that maintain a layer of air (a bad conductor of heat) next to the skin. A layer of fat under the skin serves a similar function since fat is a bad conductor of heat.

Convection is the transfer of heat along with *moving* gases or liquids, and this process accompanies conduction in removing heat from the skin. Cool air in contact with the skin is warmed by conduction; it expands, becomes 'lighter', and therefore rises upwards. Cooler air flows in to take the place of this warmed air, producing a *convection current* of air that conveys heat away from the body. Cooling by convection is increased by exposure

to moving air, e.g. to the breeze produced by a fan, since air in contact with the skin (and thus warmed by conduction) is swept away more rapidly, carrying more heat with it. Heat loss by conduction and convection is greater when the external air temperature is low. As the air-temperature rises, less and less heat is lost by these processes, until when the air-temperature exceeds the skin-temperature there is no further heat loss by conduction and convection.

We have seen that heat is lost mainly from the skin surface. But this heat was produced in the deeper tissues during respiration and has been brought to the surface of the body by the blood-stream. So the amount of heat brought to the surface of the body depends on the rate at which blood flows through the skin. When the muscular walls of the branch arteries that supply blood to the skin-capillaries contract, less blood will reach the skin and less heat will be brought to the surface. Conversely, if the diameter of these blood-vessels increases, more blood will flow through the skin, bringing more heat to the surface and making the skin warmer. These changes in the diameter (or *calibre*) of the blood-vessels in the skin are controlled by a part of the brain that is sensitive to small changes in the temperature of its blood-supply and which thus acts as a 'thermostat'* in regulating body-temperature.

When the external atmosphere is cooler than the body-temperature some heat is also lost from the lungs by warming the inspired air to body-temperature. The cooler and drier the air the greater is the heat lost in this way, through the lungs.

'The real test of our cooling powers comes either when the air is much hotter than our bodies, or when it is so saturated with water that the sweat cannot evaporate, and merely runs uselessly down our skin. It is amazing what great heat we can stand provided the air is dry. The first man that investigated* just how high a temperature he could stand was Mr. Blagden, who was Secretary of the Royal Society towards the end of the eighteenth century. He went into a room heated to 260° F., which is 48° above

the boiling-point of water. He stayed there for over three-quarters of an hour, during which time some meat was not merely roasted, but baked till it was quite hard. He took in a dog, which was kept in a basket so as not to burn its feet, and the dog managed to keep cool too. To show how evaporation cools things, he took in two pots of water. On one he dropped some wax, which melted, and prevented water vapour from escaping. The temperature of this pot rose much more than that of the one with a free surface.

'This experiment was done in dry air. Wet air is a very different story. In air at body-temperature that is completely saturated with water one sweats vigorously, but it is of no use. No water evaporates, and the body-temperature rises by 2° or 3° F. per hour. I have once been in saturated air at 120° F. It was terrible. Every breath heated my throat and lungs. We had clinical thermometers, and every minute we took them very quickly out of our mouths and read them in cold water. For, of course, in the air they would have gone up at a great rate. Our temperature rose nearly a degree a minute. I stayed in for a very short time.'[1]

FOOD AND ENERGY

In chemistry lessons we learn that the general principle underlying all chemical changes is the *Law of Conservation of Matter*, i.e. *matter can be neither created nor destroyed.* In other words, the weight of everything we start with in a chemical reaction is exactly the same as the weight of all the products formed during the reaction. In physics lessons we learn that the general principle running through the whole of physics is the *Law of Conservation of Energy*, i.e. *energy can be neither created nor destroyed but merely transformed from one form into another*, the *total* amount of energy remaining unchanged throughout the transformation.

These two laws apply equally to all living processes taking place in plants and animals. In the case of animals the truth of this has been verified experimentally by comparing the energy an animal produces in a given period with the amount of energy available

[1] From J. B. S. Haldane, *Keeping Cool.*

in the animal's food over the same period. Animals have been given a measured amount of food and kept in specially designed *animal calorimeters* for several hours, and the amount of heat-energy produced has been carefully measured (see Fig. 52). The

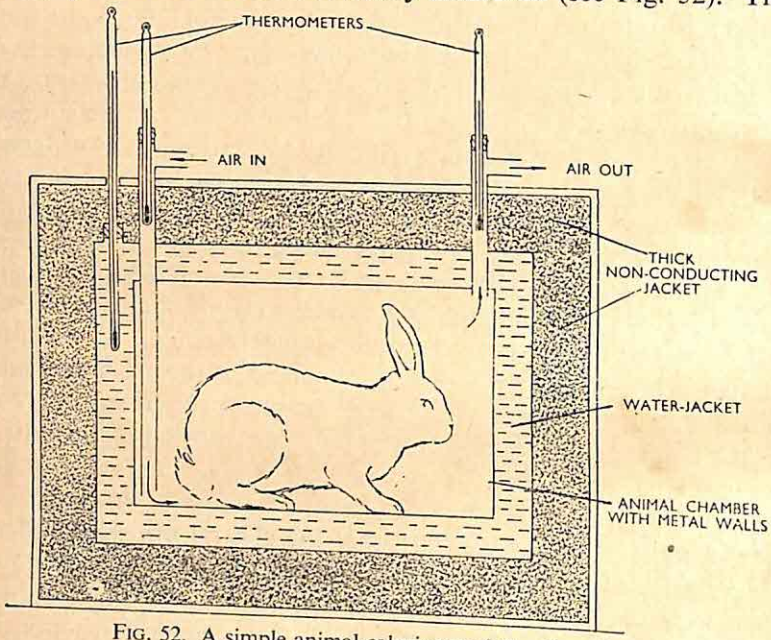


FIG. 52. A simple animal-calorimeter (after Campbell).

same amount of the same kind of food has then been burnt in oxygen in another type of calorimeter and its *heat of combustion* has been measured. In such experiments (provided that the animal neither gains nor loses weight while it is in the calorimeter) it is always found that *the same amount of food produces the same amount of heat-energy whether it is burnt rapidly in oxygen or oxidized slowly in an animal's body*. An animal, therefore, is just as much a *heat-engine* as a petrol-engine is. A motor-engine gets its energy from the rapid oxidation of a hydrocarbon (petrol), while the animal gets its energy from the slow oxidation of glucose in its body-cells with the aid of enzymes.

Since the Law of Conservation of Energy applies to animals, if we know how much energy an animal produces (as heat and work) in a day, we can calculate out how much food it must eat in order to supply this amount of energy. A man's energy-output depends on the kind of physical activity in which he is engaged, on the size of his body, and on the climate around him. It is important to realize, however, that some muscles are always working, even when a man is 'resting'; for example, his heart-muscle and the muscles concerned with breathing-movements are always *doing work*, even during sleep.

When lying down, completely relaxed and at rest (but still awake), *a man produces, every hour, about* 1 *Calorie*[1] *of heat per kilogram of body-weight.* (It is easy to remember *one Calorie per kilogram per hour.*) So if you weigh 50 kg. (= 110 lb. or nearly 8 stone) you will produce about 1,200 Calories of heat while resting completely for 24 hours. This will be your *basic energy requirement.* Any additional activity will require additional energy, e.g. merely sitting upright will increase this basic requirement by about 40 per cent.; standing straight up will need an increase of 60 per cent., as shown below.

Kind of Activity	Approximate Energy Requirement per kg. of body-weight per hour (Cal./kg./hr.)	Approximate Energy Requirement per hour for men weighing—	
		50 kg. (= 110 lb.) Calories	70 kg. (= 154 lb.) Calories
At rest, lying down	1·0	50	70
Sitting upright .	+0·4 1·4	70	98
Standing straight up	+0·6 1·6	80	112
Slow walking .	+1·0 2·0	100	140
Moderate exercise	+3·0 4·0	200	280
Severe exercise .	+5·5 6·5	325	455
Sleeping . .	−0·1 0·9	45	63

[1] The 'small calorie' that we use in our physics lessons is inconveniently small when we have to deal with the heat-production of large animals; hence the 'large Calorie' or *kilocalorie* is used. The kilocalorie is the quantity of heat required to raise the temperature of 1 kilogram of water through 1° C., i.e. 1 kilocalorie = 1,000 'small calories'.

Hence, if you know how long you spend in various kinds of bodily activity during the 24 hours, you can calculate how many energy-units (or Calories) you need from your daily food. Here are two typical examples:

Kind of Activity	Duration in hours	Calories required by 50 kg./110 lb. Student
Sleeping	9	$9 \times 0{\cdot}9 \times 50 = 405$
Sitting	8	$8 \times 1{\cdot}4 \times 50 = 560$
Standing	2	$2 \times 1{\cdot}6 \times 50 = 160$
Walking	2	$2 \times 2{\cdot}0 \times 50 = 200$
Moderate exercise	2	$2 \times 4{\cdot}0 \times 50 = 400$
Severe exercise	1	$1 \times 6{\cdot}5 \times 50 = 325$
	24	Daily requirement = 2,050

Kind of Activity	Duration in hours	Calories required by 70 kg./154 lb. Labourer
Sleeping	8	$8 \times 0{\cdot}9 \times 70 = 504$
Sitting	4	$4 \times 1{\cdot}4 \times 70 = 392$
Standing	4	$4 \times 1{\cdot}6 \times 70 = 448$
Walking	2	$2 \times 2{\cdot}0 \times 70 = 280$
Moderate exercise	2	$2 \times 4{\cdot}0 \times 70 = 560$
Severe exercise	4	$4 \times 6{\cdot}5 \times 70 = 1{,}820$
	24	Daily requirement = 4,004

These two examples will help you to understand how energy-outputs (and therefore food requirements) differ widely for different individuals, depending on their body-weight and on their occupation. Individuals also differ in the efficiency of their digestive processes, some people absorbing a larger proportion of their food than others. (The 'average' figures given on p. 152 include an extra allowance of 10 per cent. for that part of the food which is not digested or absorbed.) Climate is another important factor since in warm climates less energy (roughly 10 per cent. of the total requirement) is needed to maintain body-temperature. Human females, also, utilize a larger proportion of their available energy than do males. On the average, a woman needs only about 83

per cent. as many Calories as a man of the same weight doing similar work, i.e. about one-sixth less. In view of these very wide variations in the energy-outputs and food requirements of different individuals, any figures for 'the average man's' requirements can be only a very rough general guide. If you bear this in mind when considering the recommendations of nutrition experts and advisory bodies, then you will be able to adjust these 'average' figures to fit special classes and individual cases.

BODY-WEIGHT AND FOOD REQUIREMENTS

The simplest test of satisfactory nutrition in an adult is whether the body-weight remains steady. Too few Calories of food-energy will result either in loss of weight or in inability to do hard work. Too many Calories, efficiently absorbed, will cause an increase in body-weight, the extra energy-yielding food being stored mainly as fat. The human body is very *adaptable* and adjusts itself very efficiently to changes of diet, e.g. when Calorie-intake is less than energy-output over a long period the body loses weight until the two balance. In such cases body-fat is used up before muscle, and when muscle has to be utilized for this purpose it is the less important muscles that lose most weight, e.g. the biceps and calf muscles become greatly reduced before the heart-muscle begins to lose weight. So loss of weight under conditions of prolonged* food-shortage is usually the body's natural adjustment to the changed conditions and not the result of disease. On the other hand, marked loss of weight in a person well supplied with food is usually a danger-signal calling for investigation by a doctor. Excessive fatness, when not obviously due to habitual over-eating, may also call for medical advice.

ENERGY VALUES OF FOODS

We shall now consider the 'energy-values' of the main kinds of foodstuffs. Of the six chemical classes of foodstuffs—carbohydrates, fats, proteins, water, salts, and vitamins—only the first

three can supply energy on oxidation. The approximate energy-values of carbohydrates, fats, and proteins are:

One gram of a carbohydrate yields about 4 *Calories on complete oxidation.*
One gram of a fat yields about 9 *Calories on complete oxidation.*
One gram of a protein yields about 4 *Calories when partly oxidized.*

It is clear that the daily requirement of Calories of food-energy could be made up from many different proportions of carbohydrate, fat, and protein. For example, a well-balanced diet for a cool climate might contain

500 gm. of *carbohydrate*,	yielding	2,000 Calories,
100 gm. of *fat*,	yielding	900 Calories,
100 gm. of *protein*,	yielding	400 Calories,

making a total of 3,300 Calories (which is about the daily requirement for a 70 kg. man doing six hours of moderate muscular work in a cool climate).

Food tables[1] show the relative proportion of the various constituents in everyday foodstuffs as determined by *chemical analysis*; also their relative *calorific value*. With the help of such food-tables it is possible to calculate the food-value of any particular diet. The following example has been worked out in this way.[2]

Rice	10 ounces	(284 gm.)	1,012 Calories
Millet	5 „	(142 gm.)	455 „
Milk	8 „	(227 gm.)	136 „
Pulses	3 „	(85 gm.)	224 „
Non-leafy Vegetables	6 „	(170 gm.)	236 „
Green Leafy Vegetables	4 „	(113 gm.)	36 „
Fats and Oils	2 „	(57 gm.)	475 „
Fruits	2 „	(57 gm.)	26 „
			2,600

[1] For example, B. S. Platt, *Tables of Representative Values of Foods Commonly Used in Tropical Countries* (Her Majesty's Stationery Office).

[2] Health Bulletin No. 23, *The Nutritive Value of Indian Foods and the Planning of Satisfactory Diets*, prepared by the Nutrition Research Laboratories, Coonoor, and published by the Manager of Publications, Delhi.

This is a good example of *a well-balanced diet* supplying enough energy for a 60 kg. man doing six hours of 'moderate work' in a warm climate.

It has been suggested by nutrition experts that the relative proportions of the various foodstuffs in the diet should be such that *about two-thirds of the energy-requirement is provided by carbohydrates, one-sixth by fats, and one sixth by proteins* (*at least one-fifth of this protein being animal protein and the rest vegetable protein*), but these proportions can be varied somewhat to suit special conditions and the special needs of different people, as long as the following principles are kept in mind:

(*a*) *There must always be enough protein for growth and repair.*[1]
(*b*) *Hard muscular work requires an increased proportion of carbohydrates.*
(*c*) *Under cold conditions, an increased proportion of fat will provide extra heat.*

Human food and nutrition are discussed in greater detail in Book III A.

RESPIRATORY MECHANISM IN GREEN PLANTS

(*a*) *In the land plant*

In the leaves, stems, and roots of land plants, oxygen reaches the living cells by *diffusion* from the surrounding air. Carbon dioxide, which is produced in the cells, escapes by the same process of diffusion. At least, this is true of colourless parts and of green parts *at night*, but in the day-time the respiration of plants is hidden by the process of photosynthesis. The two gases, oxygen and carbon dioxide, pass in and out through the stomates in the leaves and green stems. Underneath these pores are air-spaces, which communicate with others extending between the cells to all parts of the plant. Plants usually respire slowly and diffusion

[1] One gram of dry protein per kilogram of body-weight, i.e. one thousandth part of the body-weight, per day, is a rough, but safe, guide for adults; growing children, however, need more than this, e.g. up to 2 gm. per kg. of body-weight.

goes on quite fast enough to supply each cell with sufficient oxygen and to remove the carbon dioxide formed.

In the stems and roots of *woody* plants, however, the stomates are replaced by larger openings, called *lenticels*, which allow diffusion of gases through the layers of cork that cover these brown surfaces. (You can see these long lenticels in any piece of bottle-cork; they run at right angles to the length of a cork, which therefore makes an air-tight stopper.)

In green leaves, in daylight, the process of photosynthesis absorbs all the carbon dioxide produced during respiration, as well as much more that diffuses *into* the leaf from the air. The same process (photosynthesis) supplies all the oxygen needed for respiration in the leaf cells in daylight.

(*b*) *In submerged water plants*

In land plants the gases have to pass through the wet cell-walls after they have dissolved in the water in the cell-wall. Submerged water plants have no stomates and gas exchange during respiration (and during photosynthesis) goes on between the leaf-skin cells (which usually lack the gas-proof cuticle of the land leaf) and the water in which the plant lives. That is, these plants get their oxygen from that dissolved in the surrounding water and they also give up carbon dioxide to the surrounding water. We know that the *illuminated* green water plant is carrying out photosynthesis as well as respiration and the net result is that it takes carbon dioxide from the water and liberates oxygen gas. The latter gas often comes out of solution and appears on the leaves as bubbles that help to make the plant float in water.

It is clear that plants do not *breathe* in the way a mammal does; also that the gas exchanges due to respiration are hidden in green plants during daylight by the reverse exchanges that accompany the food-making process of photosynthesis.

RESPIRATORY MECHANISM IN NON-GREEN PLANTS

In many fungi (e.g. moulds and mushrooms) oxygen diffuses

through the surface layer, and internal respiration is the same as in green plants. In yeast, if free oxygen is available, the same type of *aerobic respiration*† takes place; but, in the absence of free oxygen gas, yeast liberates energy from glucose by a different chemical change—*fermentation.* (In this change alcohol and carbon dioxide are the waste products instead of carbon dioxide and water.) Such a process, in which energy is liberated *without free oxygen taking part* in the change, is called *anaerobic respiration.*†

EXPERIMENTS ON RESPIRATION

(i) *To compare Inspired Air, Expired Air, and Air from the Air-sacs*

Take three similar gas-jars, A, B, and C. Fill B and C with water and invert them in a gas-trough containing a shallow layer of water. Using a piece of rubber tube, fill B with *expired air* by breathing out normally through the tube. Fill C with air from the *air-sacs* of the lungs by holding the nose and breathing out through the tube *at the end of a forced expiration.* Then invert the three jars, in turn, over a lighted candle, noting the length of time that the candle continues to burn in each case. Repeat the experiment.

	(*a*) Fresh Air	(*b*) Expired Air	(*c*) Air-sac Air
Experiment I			
Flame extinguished in	.....sec.	.....sec.	.....sec.
Experiment II			
Flame extinguished in	.....sec.	.....sec.	.....sec.
Experiment III			
Flame extinguished in	.....sec.	.....sec.	.....sec.
Average of I, II, and III	.....sec.	.....sec.	.....sec.

What do your results suggest? .....

.....

.....

(ii) *To analyse Inspired Air, Expired Air, and Air from the Air-sacs*

Take a burette fitted with a rubber tube and a spring-clip; pour

50 c.cm. of water into it and record the level. Then fill to the top with water, close the clip, and invert in a trough of water.

Open the tap and admit 50 c.cm. of fresh (inspired) air, taking care that the water-levels inside and outside the burette are approximately the same. Close the clip and put a well-fitting rubber cork in the open end of the burette—under water. Holding the burette the right way up, quickly remove the cork and drop in 5 pellets* of sodium hydroxide; then replace the cork. Shake the burette until all the sodium hydroxide has dissolved, then invert the burette in water and remove the cork. Cool the air to its original temperature by pouring water over the outside of the burette, and then adjust the water-levels as before, making sure before reading off the volume of the air that all the liquid has drained down from inside the narrow tube just below the clip. Record the new volume. (Notice that there is very little change in volume, since fresh atmospheric air contains so very little carbon dioxide—0·03 per cent.) Cork the burette again, under water, and then, holding it the right way up, quickly take out the cork and put in about 0·1 gm. of alkaline *pyrogallol*.† Replace the cork and shake the burette so as to absorb all the oxygen. (The 'alkaline pyrogallol' turns brown as it absorbs the oxygen.) Then invert the burette and remove the cork *under water*. Notice how the water rushes in to take the place of the oxygen that has been absorbed. After cooling the gas to its original temperature and equalizing the pressure, read off the new volume and record on p. 157. The gas that remains is nitrogen.

Repeat the experiment, first putting 50 c.cm. of *expired air* into the burette by breathing out *normally* through a short rubber tube attached to the burette. (Put in more than 50 c.cm. at first, cool the air down to room temperature by pouring water down the outside of the burette, then allow any excess of expired air to escape by opening the clip and adjusting the water-level.)

Again repeat the experiment, first putting 50 c.cm. of air from the *air-sacs* into the burette by holding the nose and breathing into the burette *at the end of a forced expiration.*

	Fresh Air	Expired Air	Air-sac Air
Original volume of air	50·0 c.cm.	50·0 c.cm.	50·0 c.cm.
Burette reading at start of experiment	c.cm.	c.cm.	c.cm.
Burette reading after shaking with sodium hydroxide	c.cm.	c.cm.	c.cm.
Therefore, volume of carbon dioxide absorbed =	c.cm.	c.cm.	c.cm.
Therefore, percentage of carbon dioxide, by volume =	%	%	%
Burette reading after shaking with alkaline pyrogallol	c.cm.	c.cm.	c.cm.
Therefore volume of oxygen absorbed =	c.cm.	c.cm.	c.cm.
Therefore, percentage of oxygen, by volume =	%	%	%
Therefore, percentage of nitrogen, by volume =	%	%	%

Notice that the percentage of carbon dioxide in expired air is greater than that in fresh air, while the percentage of oxygen is less. The percentage of nitrogen, however, remains about the same. Notice, also, that air from the air-sacs contains still more carbon dioxide and still less oxygen, though the percentage of nitrogen remains almost unchanged.

(iii) *To show that a small animal gives out Carbon Dioxide*

Fit up the apparatus as shown in Fig. 53. Air is drawn through the apparatus from left to right, first through soda-lime in the absorption-tube A, so as to free the air from carbon dioxide before it passes over the animal (e.g. a snail, earth-worm, frog, or cock-roach) in C. (The test-tube B, between A and C, containing lime-water, enables you to see that this air is free from carbon dioxide.) After passing through C, the air passes through the lime-water in tube D. Although the lime-water in B remains clear, that in D turns chalky, showing that the animal is giving out carbon dioxide.

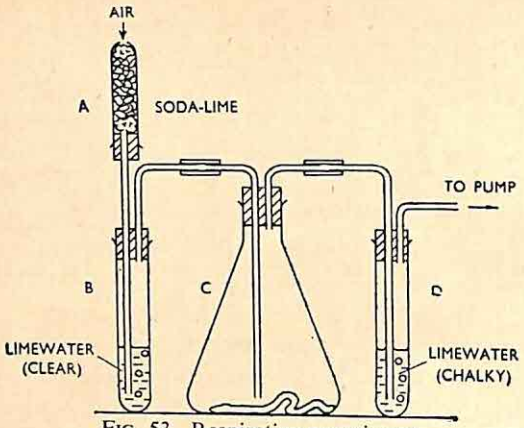


FIG. 53. Respiration experiment.

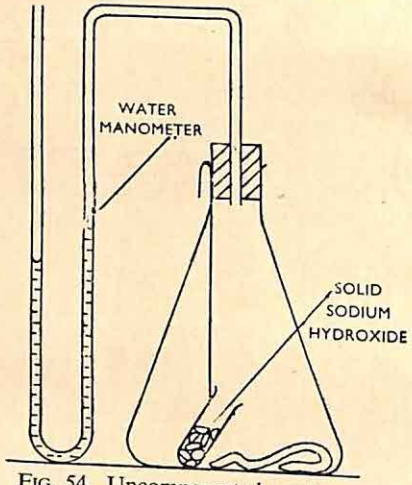


FIG. 54. Uncompensated respirometer.

(iv) *To measure the Oxygen used by a small animal during Respiration*

If a small animal is kept in a closed vessel connected to a manometer, small changes in pressure or volume can be observed. If the carbon dioxide produced by the animal is absorbed by sodium

hydroxide, the total volume of air in the apparatus will diminish as the animal uses up oxygen, and this change in volume is shown by the movement of the liquid in the manometer. In practice, however, such a simple arrangement as that shown in Fig. 54 (and

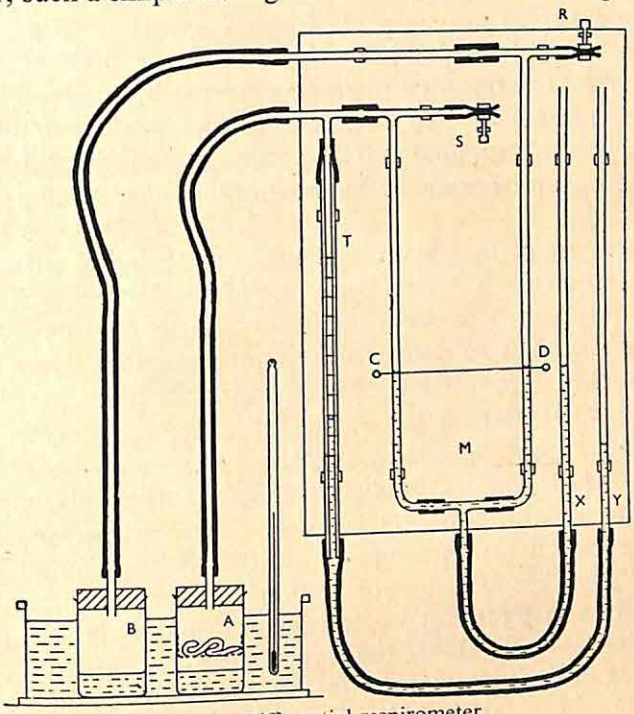


FIG. 55. Differential respirometer.

described in some books) is not satisfactory, since changes in the temperature and pressure of the air in the apparatus may produce bigger changes in volume than those produced by the animal's respiration.

The modified *Haldane apparatus*† shown in Fig. 55 overcomes this difficulty and enables us to measure the rate of respiration accurately under different conditions, e.g. at different temperatures. The two vessels, A and B (which may be either bottles, flasks, or wide tubes), are of the same size *so that any changes in tempera-*

ture and pressure have exactly the same effect on the air in A and B. Equal numbers of sodium hydroxide pellets (or equal volumes of the same sodium hydroxide solution) are placed in A and B. The animal (e.g. a small frog, snail, earth-worm, or insect) is placed in A, while B serves as a *control.* Both A and B are connected by thick-walled rubber tubing to either side of the differential water-manometer M, whose level can be adjusted by the levelling-tube X. The vessel A is also connected to the graduated tube T (a 1 c.cm. pipette graduated in 0·01 c.cm.) containing coloured water whose level can be adjusted by moving the other levelling-tube Y up or down.

At the start of the experiment the clips R and S are left open and both the vessels A and B are placed in a large vessel of water so as to keep their temperature steady. After allowing sufficient time for A and B to come to the temperature of the water, the water-level in T is set at the lowest mark and the water-levels in the differential manometer M are set at some convenient point, where the level is marked by stretching a piece of thread CD between two lumps of soft wax (or plasticene) stuck on the board to which the apparatus is attached. The two spring-clips R and S are then closed (at the same instant). As the animal uses up oxygen from A and gives out carbon dioxide (which is absorbed by the sodium hydroxide), the pressure of the air in A becomes less than that in B. Hence the water in the differential manometer M stands at a different level in the two limbs, rising in tube C and falling in tube D. The volume of oxygen used up by the animal is measured by raising the levelling-tube Y and displacing air from T until the water in the manometer M returns to its original level (also moving the levelling-tube X if the water-surfaces are above or below CD when levelled). A and B are then both at the original atmospheric pressure, and the rise of liquid in the graduated tube T gives the volume of oxygen used up.[1]

[1] This apparatus can also be used for measuring the rate of respiration of plant material. It is interesting to compare the rate of respiration of equal weights of plant and animal material, e.g. calculate the oxygen used up in unit time by 100 gm. of germinating seeds and 100 gm. of frog.

(v) *To show the effect of Temperature on the Rate of Respiration of 'Cold-blooded' Animals*

Use the apparatus shown in Fig. 55 to measure the rate of respiration of a small 'cold-blooded' animal at different temperatures. Set up the apparatus as in the last experiment, but with the vessels A and B in a water-bath whose temperature can be varied. Carry out three experiments, e.g. one at 10° C., another at 20° C., and another at 30° C. Level off the manometer-tubes and record the volume of oxygen used up every 5 minutes, continuing each experiment until three successive measurements are in close agreement. (You may find that the amount of oxygen used up is abnormal at first until the animal settles down to its new surroundings.) Take the average of the first three steady readings at each temperature.

VOLUME OF OXYGEN USED UP

Duration of Experiment	At °C.	At °C.	At °C.
0 min.	0·00 c.cm.	0·00 c.cm.	0·00 c.cm.
5 min.			
10 min.			
15 min.			
20 min.			
25 min.			
30 min.			
Average volume when steady			

(If the animal moves about much, the amount of oxygen used up will be increased by the muscular exertion. Significant results can only be obtained if the animal is at rest. This, of course, is one reason why experiments with animals are more difficult to carry out than experiments with plants—plants keep reasonably still; animals do not.)

(vi) *To show the effect of Exercise on Carbon Dioxide Production in Man*[1]

The apparatus shown in Fig. 56 consists of a mercury manometer, a No. 4 football bladder, a piece of rubber tube fitted with

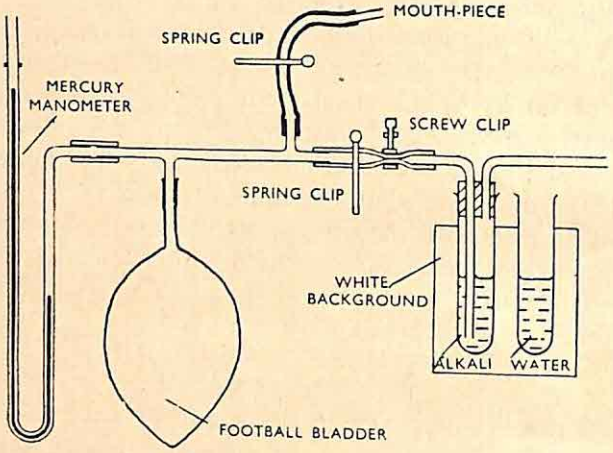


FIG. 56. Chemical respirometer.

a *spring*-clip, and a tube containing 25 c.cm. of N/500 sodium hydroxide solution with phenolphthalein as an indicator.

The horizontal piece of pressure-tubing is fitted with a *spring*-clip *and* a *screw*-clip, the latter being first adjusted to give a suitable rate of bubbling and left in this position during all the experiments. The bladder is pressed out flat between experiments.

Before taking any additional exercise, open the spring-clip on the mouth-tube and blow up the bladder *with one breath*. Then close the clip on the mouth-tube. The surface of the mercury in the open limb of the manometer will now stand higher than that in the other limb. Mark this level in the open limb by means of a rubber band. Then determine your pulse-rate, i.e. the number of times your pulse 'beats' per minute, and record this on p. 163. Then

[1] This experiment was devised by G. N. Ridley and described in *The School Science Review* for March 1936.

open the spring-clip on the horizontal tube for a few seconds so as to sweep out the air from the tubes of the apparatus before attaching the test-tube containing the alkali. Then close the clip and put the test-tube in position as shown in Fig. 56. Open the spring-clip on the horizontal tube, noting the time, and allow the expired air from the bladder to bubble through the dilute alkali and indicator, noting the time when the last trace of colour disappears.[1] Record your result below.

Squeeze all the air out of the bladder and then blow it up again after 30 seconds of vigorous exercise, e.g. 'knees up', taking care that the mercury in the manometer stands at the same level as before. Find your pulse-rate. Sweep out the air from the tubes of the apparatus as before, and then place another 25 c.cm. of the coloured alkali solution in the test-tube and take the time required to decolorize the liquid as before. Record your result below.

Repeat the experiment after 1 minute of vigorous exercise and record the 'colour-change time' and your pulse-rate below.

Duration of Exercise	Colour-change Time	Pulse-rate
0 sec.	sec.	per min.
30 sec.	sec.	per min.
60 sec.	sec.	per min.
90 sec.	sec.	per min.

Draw graphs, (*a*) plotting 'colour-change time' against time of duration* of exercise, (*b*) plotting pulse-rate against time of duration of exercise. Notice the relation between the two graphs, showing that heart-beat (pulse-rate) and carbon dioxide production are both affected in the same way by exercise.

BLOOD

THE CLOTTING OF BLOOD

Although blood is a liquid when in the blood-vessels, after escaping it soon sets to a jelly-like *clot*. Thus, if blood is collected

[1] A control tube containing water will help in judging the colour change.

in a beaker and is then allowed to stand for 5–10 minutes, the beaker can be inverted without any blood running out. The blood clots, owing to the formation of a network of fine threads of *fibrin*,† an insoluble protein, formed from *fibrinogen*,† a soluble protein contained in the blood, when blood is exposed to the air. This net of fibrin threads weaves itself around the blood-cells. After standing for some time the fibrin threads contract so that the clot contracts also, squeezing out a clear yellow liquid called *serum*.† When blood escapes from a wound it begins to clot and, if the flow of blood is not too rapid, the clot soon plugs up the wound and forms a foundation network for the growth of new tissue to repair the damage. In this way, clotting is Nature's 'first aid' to the healing of wounds.

If freshly-collected blood is stirred or shaken, the fibrin can be removed as it is formed and the *de-fibrinated blood* that remains will not clot.

Since the calcium salts in the blood are necessary for clotting to take place, if the calcium compounds are removed the blood will not clot. Potassium oxalate† or ammonium oxalate† will precipitate the calcium from the blood as insoluble calcium oxalate.† This *oxalated blood* can be made to clot by adding a calcium salt.

(i) Place strips of red and blue litmus-paper on a clean white tile and wet them with drops of 10 per cent. sodium chloride solution. Place one drop of blood (either de-fibrinated, oxalated, or fresh—your own, for example) on each strip, and after about 30 seconds wash off the blood with more salt solution. Notice that the blood is slightly alkaline to litmus.[1]

(ii) Examine a drop of blood under the microscope, first spreading the drop with a second glass slide so as to form a thin film of

[1] Drops of your own blood can best be obtained as follows: Sterilize* the skin at the tip of the left fore-finger by wiping with cotton-wool soaked in alcohol. Then wrap the corner of your handkerchief, or a piece of string, several times round the end joint of this finger (but do not *tie* it) and bend the finger so that the finger-tip becomes swollen with blood. With a sterile needle (which has been kept in alcohol or passed through a flame) pierce the skin sharply and squeeze out a drop of blood. Wipe the spot with alcohol after removing the blood. This method is safe and practically painless.

blood. Using the high-power objective, identify the red blood-cells (reddish-yellow under the microscope) surrounded by colourless *plasma*.

(iii) Take about 25 c.cm. of oxalated blood that has been warmed to about 37° C. and add 1 c.cm. (25 drops) of 2 per cent. calcium chloride solution. Mix and then allow to stand. Note the formation of a *blood-clot*. After standing for about an hour, notice how the clot begins to contract and becomes surrounded by *serum*. Examine this blood-clot. Notice the bright red colour on the surface that is in contact with the air. Cut the clot in half and notice that the part which has not been in contact with the air is a dark red colour, and that it turns bright red on standing in contact with the air.

(iv) Dilute 10 c.cm. of blood to 100 c.cm. with water. Fill two 5-in. by $\frac{5}{8}$-in. test-tubes with this diluted blood (20 c.cm.) and to one tube add 1 c.cm. of 1 per cent. *sodium hydro-sulphite*† *solution* ($Na_2S_2O_4$)—a weak *reducing agent*. (This solution must be freshly prepared.)

Close the mouth of the tube with the thumb and mix gently. Notice that the bright red colour changes to dark red because the *oxy-haemoglobin* (characteristic of *oxygenated* blood) has been reduced to *haemoglobin* (characteristic of *de-oxygenated* blood). Transfer this dark red liquid to a larger test-tube and shake vigorously in contact with air. The oxygen of the air oxidizes the haemoglobin back to oxy-haemoglobin and the bright red colour returns. Transfer to the original test-tube and compare the colour with that of the 'control' tube of diluted blood.

EXAMINATION OF HEART AND LUNGS

Examine the heart and lungs of a large mammal, e.g. sheep or goat, obtained from the abattoir with the blood-vessels cut off some distance from the heart and with the wind-pipe and voice-box complete. Slit open the *heart-sac* to expose the heart and identify (*a*) the ventral surface of the heart with its diagonal groove filled with fat, (*b*) the right ventricle (thin-walled and rather soft

to the touch), (*c*) the left ventricle (thick-walled and therefore firm and hard), (*d*) the right and left auricles (thin-walled and divided from the ventricles by a horizontal groove filled with fat). Draw the heart and lungs as seen from the ventral surface.

Cut horizontal slices across the apex of the heart until you expose the cavities of both ventricles. Notice (*a*) the shape of the cross-section of each ventricle and (*b*) the relative thickness of

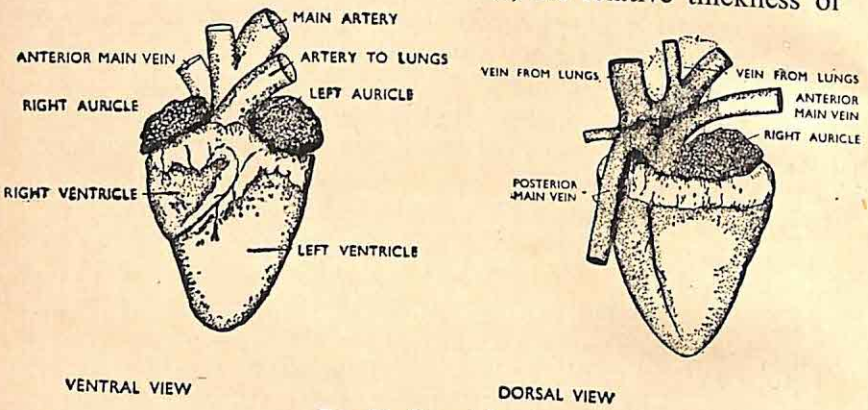


FIG. 57. Sheep's heart.

their muscular walls. Draw what you see. Wash out the heart with 10 per cent. sodium chloride solution and then slit up the ventral wall of each ventricle. Notice (*a*) the valve-strings and valve-flaps of the valves that separate the ventricles from the auricles (there are 3 valve-flaps in the right side of the heart and 2 in the left), (*b*) the opening of *the artery to the lungs* in the *right* ventricle, (*c*) the opening of *the main artery to the body* in the *left* ventricle. Push a thick, round-ended glass rod into these openings and follow the course of these two arteries.

Slit up the auricles and open up the right and left sides of the heart. Notice (*a*) the openings of the two *main veins* (anterior and posterior) into the *right* auricle, and (*b*) the openings of the two *veins from the lungs* into the *left* auricle. Follow the course of these veins by pushing a glass rod along them.

Finally, slit open the main artery just above the left ventricle to expose the three *pocket-shaped valves* that prevent back-flow of blood into the ventricle. Pour some water on to these valves and notice how they expand to meet each other and thus close the opening.

Examine the lungs and wind-pipe of a large mammal and draw their external appearance from the ventral view. Identify (*a*) the voice-box, (*b*) the *epiglottis*, a lid-like structure that closes the opening to the wind-pipe when swallowing. Look down inside the voice-box and notice the *glottis*, a narrow slit, with the *voice-cords* on either side. Slit open the voice-box and find the voice-cords.

Examine the wind-pipe and notice how it is stiffened by 'C'-shaped rings of gristle so that it is always kept open. Slit open the wind-pipe as far as its posterior end, where it divides into two branches. Slit open one of these branches and follow its course as it divides again and again into smaller and smaller air-tubes. At the same time, identify the arteries and veins running through the tissues of the lungs.

(See *General Science Workbook*, Book Two, pp. 76–81, for learning exercises on this chapter.)

Chapter VII

EXCRETION IN PLANTS AND ANIMALS

In Book Three, Chap. II, we learnt something about *excretion*, and we saw that this term is applied only to the removal of *internally produced waste products from the body-cells*. The semi-solid, undigested food-remains, passed out as *faeces*, have never been inside the body cells but have been *rejected*,* or *eliminated*,* or *evacuated*, by the body, and *not* excreted.

EXCRETION IN ANIMALS

In microscopic, one-celled animals like Amoeba and Paramecium, waste products can be excreted simply *by diffusion* through the surface layer, or they can be squeezed out by contractions of the protoplasm. An earth-worm has special excretory organs in each segment of its body—fine coiled tubes opening through the body-wall or leading into the food-canal. Body-fluid, containing waste products, can pass along these 'excretory tubes' from the body-cavity to the outside, either through the skin or by way of the food-canal. Part of each 'excretory tube' is surrounded by a network of capillary blood-vessels, so that waste products are absorbed from the blood and are then passed out through the 'excretory tube'.

In insects (e.g. the Cockroach) the hind gut bears many yellow thread-like excretory tubes that remove waste nitrogenous matter from the blood and discharge it into the hinder part of the food-canal (see Fig. 44). Insects use some of their excreted nitrogen compounds to build up their horny, external skeleton, which is composed of *chitin*,† a nitrogen-containing polysaccharide with long fibrous molecules.

In the Snail the excretory organ or 'kidney' is a triangular body in the roof of the mantle-cavity, near the heart (see Fig. 113). It

consists of a hollow sac with spongy walls that are well supplied with capillary blood-vessels from which nitrogenous waste products are absorbed and discharged to the outside by a narrow 'kidney-duct' running alongside the large intestine.

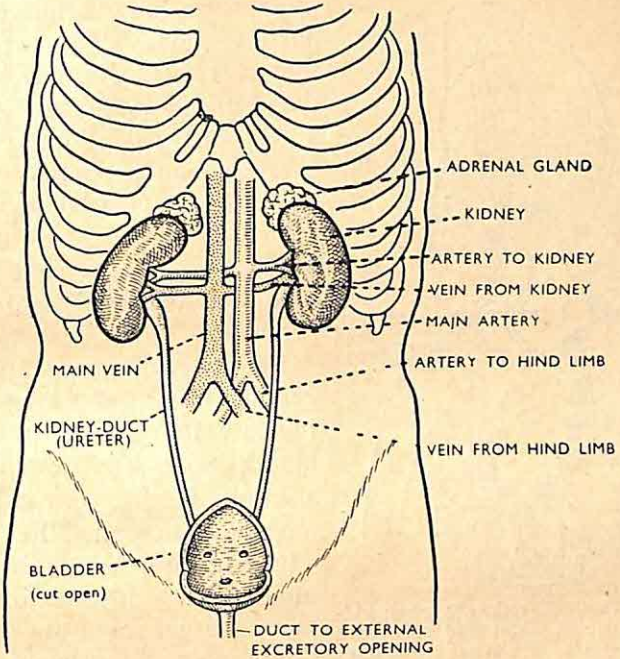


FIG. 58. Human kidneys and their blood supply (diagrammatic).

In all the backboned animals, e.g. amphibians, fishes, birds, and mammals, the main excretory organ is the *kidney*, and we shall now discuss the mammalian kidney as a fairly typical example of the kidneys found in all vertebrate animals.

The kidney of a mammal is a bean-shaped gland attached to the dorsal wall of the body-cavity. It is made up of large numbers of 'kidney-tubes' packed closely together and surrounded by a network of capillary blood-vessels. These are supplied with blood by a large *renal artery* leading directly from the main dorsal artery

(the *dorsal aorta*), and a large *renal vein* returns blood to the posterior main vein (see Figs. 58 and 59). The kidney-tubes become filled with a liquid (coming from the surrounding blood-capillaries), which contains urea and some other soluble salts, and these tubes all discharge into a narrow kidney-duct (or *ureter*) that carries the liquid (*urine*) to the *bladder*, where it is stored and discharged at intervals. An average man produces about $1\frac{1}{2}$ litres of urine each day—containing about 30 gm. of urea, 15 gm. of common salt, and smaller quantities of other soluble waste products.

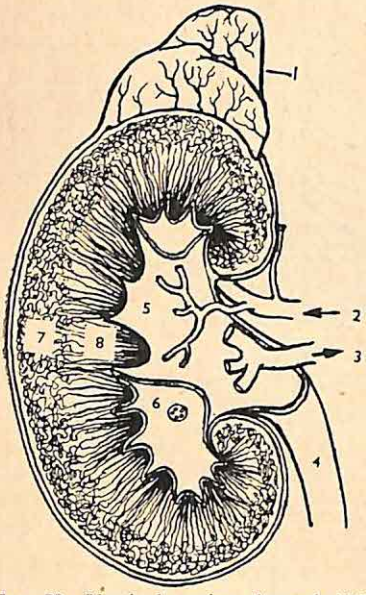


FIG. 59. Vertical section through kidney. 1. Adrenal gland (not concerned with excretion). 2. Artery to kidney (renal artery) showing only beginnings of its branches. 3. Vein from kidney (renal vein) showing only beginnings of its branches. 4. Kidney-duct (ureter) draining urine to bladder. 5. Cavity into which urine drains from kidney-tubes. 6. Conical projections into cavity through which kidney-tubes discharge urine. 7. Outer layer (cortex) containing coiled ends of kidney-tubes. 8. Inner layer (medulla) containing straight portions of kidney-tubes.

Besides getting rid of nitrogenous waste, the kidneys also regulate the composition of the blood. Soon after we have drunk a lot of water, the blood becomes more dilute, but the kidneys adjust its composition by removing the excess of water and excreting a larger quantity of more dilute urine. Conversely, after we have eaten a lot of salt, the percentage of salt in the blood rises, but the kidneys at once begin removing more salt and less water (i.e. they excrete a more concentrated urine) until the composition of the blood returns again to normal. Similarly, when sweating is rapid and the blood is giving up water through the skin, less water is absorbed from the blood by the kidneys and a smaller volume

of more concentrated urine is excreted. Under cold conditions, when very little water is being lost as sweat, the kidneys absorb more water from the blood and excrete a larger volume of more dilute urine.

In reptiles and birds much water is absorbed from the urine on its way from the kidneys, and these animals produce a semi-solid, white excretion, which is mainly *uric acid*.† It also contains phosphorus compounds and plays an important part in the *phosphorus cycle*.

In some parts of the world, sea birds live in countless thousands. Most of their droppings accumulate near to their nesting-places on the shore, and in time they produce very deep layers of *guano*,† which forms a valuable nitrogenous manure. The rock phosphates of some Pacific islands are formed indirectly from this guano. The phosphates are slowly dissolved by rain-water from the guano and then react with the calcium carbonate of the underlying coral to form an insoluble phosphate rock, often of great thickness. The phosphorus that it contains came originally from other phosphate rocks and from dead creatures. This phosphorus was carried into the sea by rivers, taken from the sea by fish, these in turn were eaten by birds, and the phosphorus (or part of it) was excreted. So the 'superphosphate' that farmers now put on their land has had a long history.

EXAMINATION OF KIDNEY

Examine the kidney of a mammal (e.g. sheep, ox, rabbit, or goat). Identify (*a*) the *renal artery* bringing blood to the kidney, (*b*) the *renal vein* taking blood away from the kidney, (*c*) the kidney-duct (or *ureter*) carrying urine from the kidney to the bladder. With scissors, cut open the kidney-duct and trace it up to where it enlarges into a funnel-shaped cavity inside the kidney. Notice the conical projections inside this cavity—this is where the kidney-tubes discharge their urine into the cavity. Cut open the kidney longitudinally into two halves and notice that there are two different layers (see Fig. 59). The outer, dark-brown layer consists

of the coiled ends of the kidney-tubes and the closed, outer end of each tube is in close contact with a bunch of capillary blood-vessels. The inner layer is of a lighter colour and appears striped. This layer consists of the straight portions of the kidney-tubes that collect the urine and discharge it into the central cavity from whence it is drained to the bladder by the kidney-duct.

In Man and some other mammals the skin plays a small part in excretion, although the function of sweating is not primarily getting rid of waste matter but regulation of body-temperature. Sweat, however, is not pure water. It contains small quantities of mineral salts (as shown by its taste) and urea. Although there is only a very small quantity of urea in sweat, it gives a smell to sweaty clothes when it decomposes.

EXCRETION IN GREEN PLANTS

Plants excrete the waste products of respiration (carbon dioxide and water), of photosynthesis (oxygen), and surplus* water absorbed by the roots. Plants produce very little nitrogenous waste.

The carbon dioxide produced during respiration can be got rid of, from all parts of a plant, through the lenticels and stomates. In green parts of plants, in the light, the carbon dioxide formed during respiration is used for photosynthesis. In the same way some of the oxygen produced during photosynthesis is used for respiration, and only the surplus is excreted through the stomates.

A plant takes in much more water than it requires to keep its cells turgid and for synthesizing carbohydrates. This surplus water is given off as water-vapour, from the leaves, during transpiration.

Excretion in plants, therefore, is a much simpler subject than excretion in animals.

(See *General Science Workbook*, Book Two, pp. 76–81, for learning exercises on this chapter.)

Chapter VIII

MOVEMENT AND BEHAVIOUR IN LIVING THINGS

Movement is a characteristic of all living things. It is more active and noticeable in animals than in plants. As a general rule, movements (apart from those of normal growth and continuous movements like heart-beat and breathing) are made in response to a *stimulus*, i.e. to any change in the environment that affects the activities of living things, all of which are extraordinarily sensitive to changes in their environment. Usually a movement in response to a stimulus is such as to favour the individual making the response—*to help it to survive.*

PLANT MOVEMENTS

Most plants are firmly rooted in the ground and they do not move about in search of food or to avoid enemies. Nevertheless, they show marked, but slow, *growth-movements.* These are usually responses to the stimuli of *gravity*, *light*, and *water*, and their effect is to give the plant a greater chance of survival. Thus roots grow downwards into the soil (where they get water and dissolved mineral salts), stems grow upwards into the light and air, flowers and stomates open during the day, seeds and fruits are shed when they are ripe, tendrils curl around supports and hold the climbing plant firmly in place.

PLANT MOVEMENTS IN RESPONSE TO GRAVITY

In whatever position a seed is planted the main root always grows downwards and the main stem upwards. In each case the direction of growth is the indirect result of the *force of gravity.*

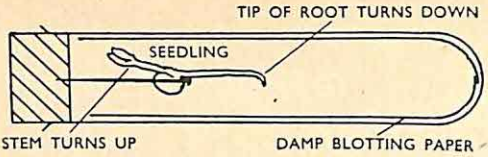


FIG. 60. Effect of gravity on direction of growth.

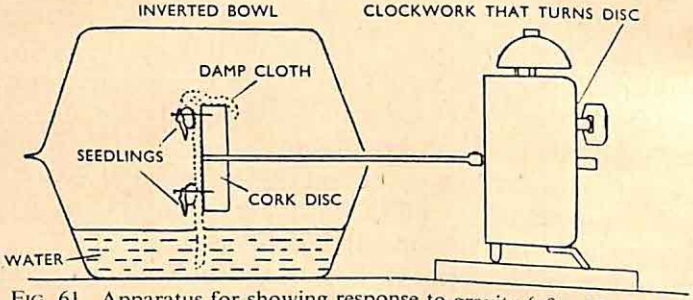


FIG. 61. Apparatus for showing response to gravity (after Hatfield).

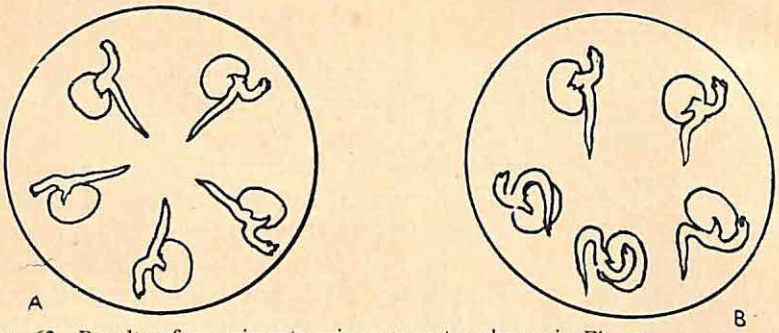


FIG. 62. Results of experiments using apparatus shown in Fig. 61. A. Seedlings grown on vertical disc kept turning slowly. B. Seedlings grown on stationary vertical disc.

To show the effect of gravity on the direction of growth by roots and shoots.

In Book One, Chap. IV, we did a simple experiment (see Fig. 60) showing that when a seedling is placed in damp air, with its

main root and shoot horizontal, the end of the root turns vertically downward. The only force that has acted on the root from below is that of *gravity*. If we make equidistant marks (with Indian ink) on the root before starting the experiment, we find that the part of the root which *turns* is the part that is growing in length. This turning movement brings the root-tip into the vertical position again. The older part of the root (which has finished growing in length) remains horizontal.

In the same experiment we saw that the main stem turns upwards, bringing the terminal bud vertical. Again, it is only the growing part of the stem that bends. These movements of root and shoot are *non-reversible growth movements*.

We cannot set up a 'control' experiment in which there is no force of gravity acting on the plant, but we can make gravity act equally in every direction by fastening seedlings to a *vertical* disc on a horizontal axle that can be turned by a clock, as shown in Fig. 61. When the disc revolves slowly (e.g. once every hour) the direction of the force of gravity relative to the seedlings is continually changing and is thus neutralized, as it were, and the seedlings continue to grow in the direction in which they were placed at the beginning of the experiment (see Fig. 62A). In a control experiment with seedlings fastened to a similar vertical disc that is *not turned*, it is found that, in response to the force of gravity, the root turns downwards and the shoot turns upwards (see Fig. 62B).

If a similar experiment is carried out with seedlings fastened to a *rapidly revolving* vertical disc (turned by a small water-turbine or electric motor), an *outward force* is set up that acts in the same way as the force of gravity, and the roots grow outwards and the shoots inwards towards the centre of the disc (see Fig. 63). If the experiment is repeated with the rapidly revolving disc in the horizontal plane, the mechanical force acts outwards and the force of gravity acts downwards. The seedlings respond to both forces, so that the roots grow outwards and downwards and the shoots inwards and upwards.

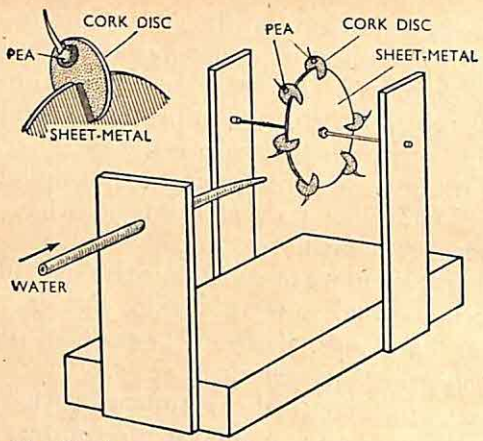


FIG. 63. Simple water-wheel for experiments on direction of growth (after Palmer).

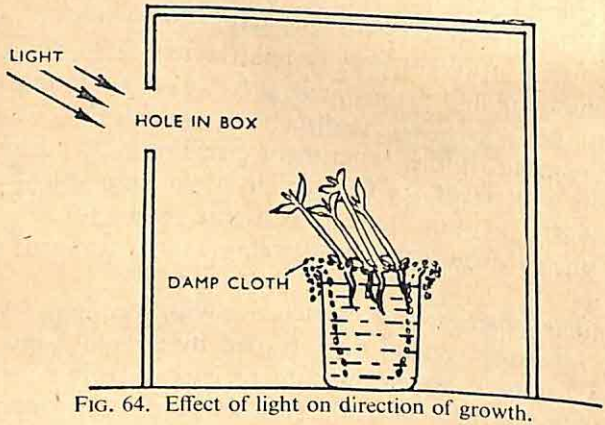


FIG. 64. Effect of light on direction of growth.

PLANT MOVEMENTS IN RESPONSE TO LIGHT

Plant stems usually grow vertically upwards and this we have found to be a response to the downward stimulus of gravity. Under normal conditions these stems are evenly illuminated. But observation and experiment show that *light* (as well as gravity) influences the direction of growth in a stem.

In Book One, Chap. IV, we studied the effect of unequal illumination on the direction of growth by stems and leaves (see Fig. 64). When the seeds germinated, the stems grew towards the hole and the green leaves grew at right angles to the beam of light entering the box. As a result, the leaves receive the maximum amount of light for photosynthesis. As with the root, marking

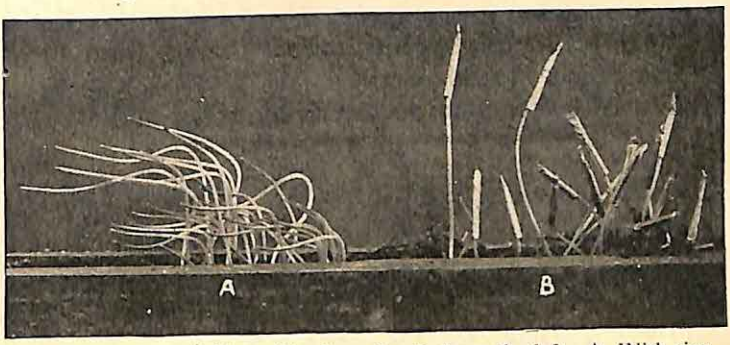


FIG. 65. Grass seedlings illuminated only from the left. A. With tips uncovered. B. With tips covered.

the shoot at equal distances shows that the curvature of the stem takes place *only in the growing region*. The leaves turn because of similar, but less obvious, growth changes in their stalks.

PLANT HORMONES

There have been many attempts to find out the way in which plants *receive* the stimuli of gravity and light. Certainly there are no *nerves* in plants and no special *sense-organs* as we know them even in the lower animals. A simple experiment will give you some idea of the way in which this difficult problem has been approached and partly solved.

Demonstration. Grow about 30 oat or millet seedlings in a shallow rectangular tray filled with *moist soil*. The seeds should first be soaked in water for 12 hours and then buried to their own depth in the soil. Place the tray *in darkness* after covering it with a box lined with wet blotting paper. After 4–5 days (the time varies with the temperature) the seeds will have germinated and

the shoot will be a simple vertical colourless tube (really a part of the seed-leaf with the true leaves growing up inside it). Examine these shoots at intervals *in red light* (such as is used in photographic dark rooms) and, when they are about 10–15 mm. long, treat them as follows:

(i) Cover the extreme tips of 10 shoots with tiny caps of tin or aluminium foil.

(ii) Remove the extreme tips (about 1 mm.) from 10 others, using a sharp razor blade.

(iii) Leave 10 shoots untouched as controls.

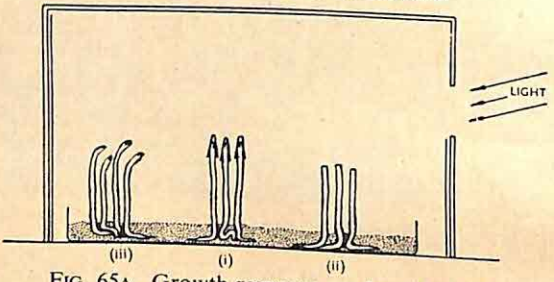


FIG. 65A. Growth response produced by light.

Now place the tray of seedlings inside another box and illuminate them *from one side only*, through a horizontal slit. Examine after 2 hours.

It will be seen that the control seedlings have all bent towards the light while *those with their tips removed or shaded have remained straight, although they have grown in length about as much as the controls* (see Figs. 65 and 65A).

This experiment shows that the shoot tip is a kind of sense-organ: it *perceives** the light stimulus and passes it down to the growing regions, which *respond.*

Many experiments of this kind have shown that a *hormone*† or 'chemical messenger' is concerned with the growth-movements of roots and shoots. This hormone (whose chemical formula is known) is produced in the extreme tip of shoots and roots and it passes to the growing regions, whose rate of growth it controls.

When the tip is exposed to light coming from one side only, the hormone travels *sideways* at first instead of straight along the shoot or root. As a result, it later brings about unequal growth on the two sides of the shoot or root, which therefore bends.

These movements are clearly to the plant's advantage. A root, in growing downwards, buries itself in the soil; while the stem, in growing towards the light, brings the leaves into the most favourable position for photosynthesis. In addition, the leaves usually arrange themselves so that they do not overlap and shade one another. This arrangement of the leaves is called a *leaf-mosaic** (see Fig. 66). When a plant receives light from one side only, the directive influence of light on the stem may be as strong as that of gravity.

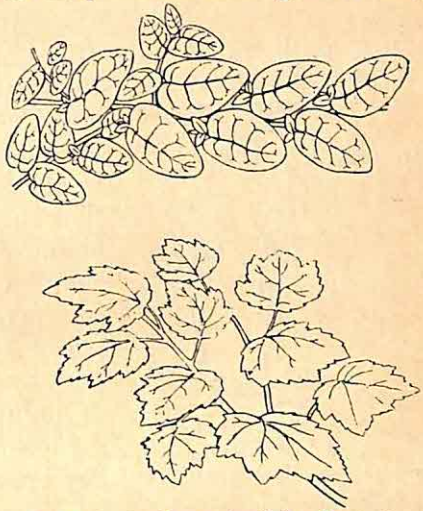


FIG. 66. Leaf-mosaics (after Fortey).

PLANT MOVEMENTS IN RESPONSE TO WATER

In Book One, Chap. IV, we also showed that roots grow towards water and that *the effect of water upon the direction of growth of the root may counterbalance that of gravity*. You will also remember that in Book Two, Chap. VI, we emphasized that, in the soil, roots grow towards water and that the water does not usually move towards the roots.

Entirely different kinds of plant movement in relation to water are carried out by shoots. The stomatal pores, for instance, are normally open in the light, but they close when the leaf *wilts* owing to lack of water. Again, many fruits split open only when the cell walls have dried out completely.

There are other plant movements that we cannot deal with

here—the opening and closing of flowers, movements of stamens, of climbing stems, and of tendrils. Most of these movements are responses to external *stimuli* and they may all be regarded as *adaptations* that help the plant to survive and to reproduce itself.

MOVEMENT IN ANIMALS

Just as there are some free-swimming plants, there are some exceptional animals that remain fastened to one spot, e.g. Sponges and Corals. By far the greater number of animals, however, move from place to place in search of food and to avoid their enemies. This movement is brought about in various ways, but even in the simplest animals it seems to involve *contraction* of some part of the protoplasm. Contraction is a process that requires energy, and this is obtained during respiration.

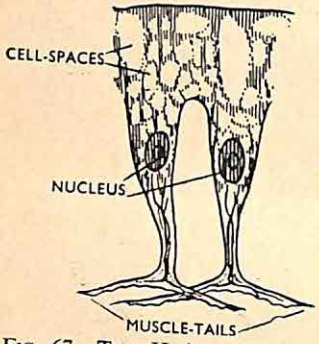


FIG. 67. Two Hydra cells with 'muscle tails' (highly magnified).

In *Amoeba* part of the protoplasm flows out to form a rounded projection in the direction in which the animal is moving. The rest of the protoplasm then flows in the same direction, so that the cell moves forwards with a rolling 'tank'-like motion over the surface on which it rests.

In *Paramecium*, another one-celled animal, the whole surface is covered with tiny protoplasmic threads, called *cilia*. These cilia carry out rapid, whip-like movements and, as a result, the animal moves through the water. As the cilia are arranged in spiral rows, they give Paramecium its characteristic spiral movement.

In *Hydra*, which consists of a large number of cells arranged in *two layers* to form a hollow sac, some of these cells have 'muscle-tails' that can contract. The 'muscle-tails' of the outer layer of cells run lengthwise up and down the animal, while the 'muscle-tails' of the inner layer run round the body. When the longitudinal 'muscle-tails' of the outer layer contract, the animal

shortens in length, and when the circular 'muscle-tails' of the inner layer contract the animal becomes long and thin. Similar contractions produce the movements of the hollow threads (tentacles) surrounding the mouth (see Fig. 67).

Although Hydra spends most of its life attached to one spot, e.g. to a piece of water-weed, it is able to move from place to place. It can change its position in two different ways. Slight

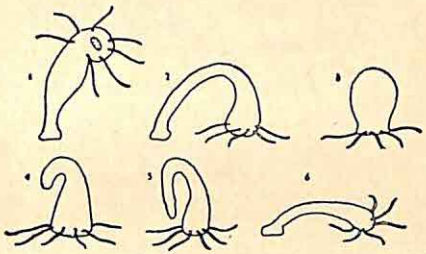


FIG. 68. 'Walking' movements of hydra.

changes of position are carried out by slowly *sliding* along on its 'foot'. More rapid movements are carried out by '*walking*' (see Fig. 68). The animal bends over and attaches its 'mouth' to the supporting surface. It then lifts its 'foot' and contracts strongly before extending its body again, then swinging its 'foot' into the required direction and attaching it once more. The 'mouth' is then lifted and the Hydra animal 'stands up' again.

In the *earthworm* the body-wall consists mainly of two layers of muscles—an outer layer of *circular muscle-fibres* running round the animal and an inner layer of *longitudinal muscle-fibres* running up and down the animal from head to tail. The earthworm's body-cavity is filled with liquid, and the pressure of this liquid keeps the body-wall stretched and firm (like a turgid plant-cell). Rings of tiny bristles project from each segment. The earthworm moves by alternate contractions of its two layers of muscles. When the circular muscles of the hind part contract, the diameter of this part of the body is reduced and the liquid in the body-cavity is

compressed, thus pushing the front part of the animal forwards. The bristles at the head end are then pressed into the soil, while the longitudinal muscles contract and draw the hind part forward.

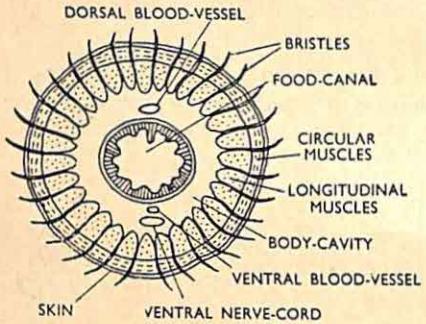


FIG. 69. Transverse section of Indian earthworm (very diagrammatic).

The bristles at the tail end are then pressed into the soil and the whole process is repeated.

The *snail* has a remarkable gliding* motion. The 'foot' is a mass of soft muscle, and alternate waves of contraction and

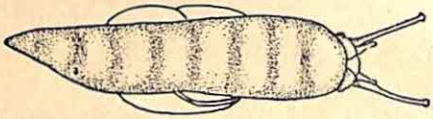


FIG. 70. Ventral view of snail moving over a sheet of glass, showing waves of muscular contraction passing along animal's 'foot'.

relaxation run forward through these muscles, carrying the animal forwards along its slimy track.

THE SKELETON AND MOVEMENT

The movements of small, soft-bodied animals are comparatively slow. The more rapid movements of larger animals are only possible because they have muscles attached to rigid supports provided by parts of the *skeleton*.

The most primitive type of supporting tissue in animals is the *external skeleton* found in many Invertebrates. Its simplest form is an unjointed shell. In many microscopic aquatic animals this shell serves only for protection and support, playing no part in movement. The shell of a snail serves a similar purpose, and the animal retires from its enemies by drawing itself back inside the shell. In insects, however, we find an entirely different type of external skeleton—a *jointed skeleton*—where the muscles of movement can pull on the jointed limbs and thus enable the animal to move from place to place. This hollow, jointed, tubular skeleton is very efficient for support and protection, and some insects are among the most successful and efficient of all animals. But a dead, inelastic, external skeleton has one very serious disadvantage—*it hinders growth.* We have seen that all insects cast off their external skeleton several times during their life-history. It is often difficult for the soft-bodied animal to get out of its horny covering, and it is helpless and defenceless until the new covering grows and hardens. We have also seen that many of the higher insects, e.g. butterflies, moths, and mosquitoes, undergo *complete metamorphosis* during their complicated life-history, so that growing and skin-casting take place in the larval stage, before the development of any complicated wings, jointed walking-legs, and jointed mouth parts.

An insect's external skeleton is composed of *chitin*, a horny substance similar to the material forming human finger-nails and hair.

In crabs and lobsters the external skeleton is thickened and hardened by the addition of *calcium carbonate* to the horny *chitin.*

In Vertebrates the supporting tissue is arranged on an entirely different plan: there is an *internal skeleton.* This bony framework is built on the same essential ground-plan in all *Vertebrates*, e.g. fish, frog, bird, and mammal, consisting of a *main axis* (*skull, backbone, ribs, breastbone*) and two pairs of *attachments* (*limbs*), which may be either legs, or arms, or fins, or wings. These are attached to the two *limb girdles* that form part of the skeleton.

An internal skeleton does not give such complete protection to the soft parts of the body as the external skeleton of an insect, but it offers no hindrance to growth and it is better adapted to free body-movements. We have already discussed the general arrangement of bones in the vertebrate skeleton (Book Three, Chap. II), but we have still to explain how movements of bones are brought about by the contraction of muscles.

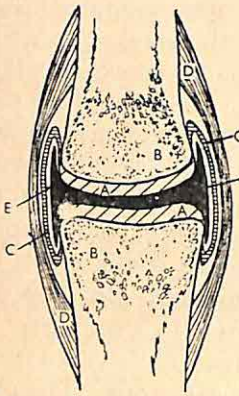


FIG. 71. Structure of freely movable joint (diagrammatic). A. Layer of gristle (cartilage). B. Bone. C. Joint-capsule containing synovial fluid. D. Binding ligament. E. Synovial membrane forming sac enclosing joint. (N.B. The cartilage-covered ends of the bones are actually in contact, but in this diagram the bones are pulled apart to show the joint-structure more clearly.)

Where neighbouring bones (e.g. most of the skull-bones and the bones of the hip-girdle) do not move independently, they often grow together. Many bones, however, are connected together by *joints* that enable two neighbouring bones to move freely when pulled by skeletal muscles. In the human body there are 230 joints, large and small —as many as 25 are present in each leg.

The moving parts of the body are almost perfectly lubricated. Fig. 71 shows the way in which a joint is constructed. The jointed bones do not actually touch one another but are bound together by strong fibrous tissue forming the 'capsule'.† They are not, therefore, easily pulled out of joint, or *dislocated*.* In joints like the hip and the shoulder the rubbing surfaces are smooth and rounded and the ends of the bones are covered with a thin layer of translucent white *cartilage*.† This is not only elastic, thus acting as a cushion, but it is also covered with a film of slimy fluid that is the actual lubricant of the joint. This *synovial fluid*† fills the closed sac that separates the two layers of cartilage at the joint. It is formed from living cartilage as this gradually wears away and is replaced by new growth.

The type of joint that allows the freest movement in all direc-

tions is the '*ball-and-socket*' type, which joins the limbs to the limb-girdles (hip- and shoulder-joints) (see Fig. 72). '*Hinge-joints*' (e.g. those of the elbow, knee, jaw, and finger) allow movement in one plane only. The joints between the individual bones (*vertebrae*) of the back-bone are so firmly bound together by tendons that

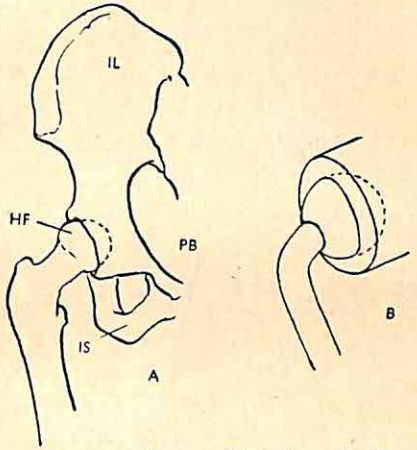


FIG. 72. A. A human hip-joint. B. Simple mechanical ball-and-socket joint. HF. Rounded head of thigh-bone (femur). IL, PB, IS. Fused bones of hip-girdle (after Thomson).

very little movement can take place, since the main function of the backbone is to provide strength and support.

The muscles that move the bony levers of the body are, of course, living. They are made up of cells that have the special property of being able to *contract*. They can exert a pull but not a push. So we find that every voluntary muscle is attached to at least two jointed bones, and that nearly all these muscles are arranged in pairs. One muscle brings the jointed bones closer together when it contracts, the other pulls them back again to their original position. Nervous control of muscular contraction is such that when one member of a muscle-pair contracts, the other relaxes. In Fig. 159 is shown only one of the muscle-pair in each example, together with a diagram of the type of lever concerned.

Fig. 73 shows two paired muscles of the arm; the *biceps* muscle raises the fore-arm, while the *triceps* lowers it. The mechanical advantage (see p. 306) of the 'biceps lever' is only about $\frac{1}{10}$, but

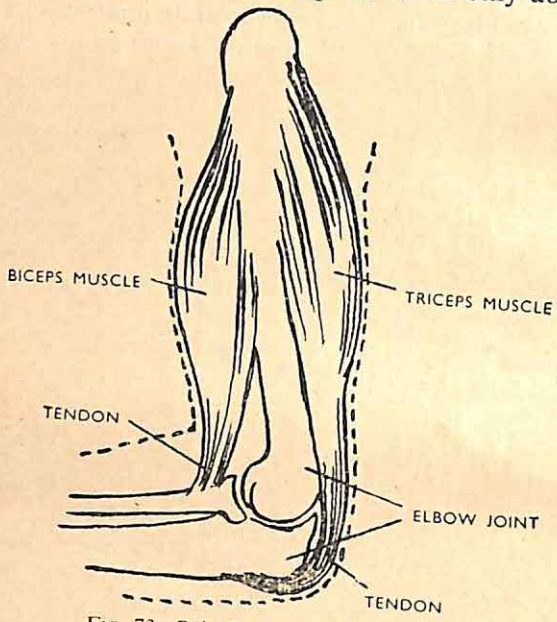


FIG. 73. Paired muscles of upper arm.

it has other advantages, e.g. it enables the hand to move *rapidly* through a wide angle while the muscle makes only a small contraction.

'SKIN-SKELETONS'

The internal skeleton of a Vertebrate gives less effective *protection* than the external skeleton of an insect. In most Vertebrates, however, additional protection is given by a '*skin skeleton*' covering the body. Fishes and reptiles have their skin covered with *scales*; birds have *feathers*; mammals have *hair*. All these are external protective coverings. In gristly fishes, e.g. sharks, the skin is covered with thousands of tiny scales. Each scale arises

from a flat, bony base embedded in the skin and consists of a backward-pointing spine of *dentine*, covered with a layer of *enamel*. There is a *pulp-cavity* in the centre of the spine. These scales of primitive fishes are of exactly the same structure as the fish's *teeth*. They are also very much like our own teeth in structure (see Fig. 74).

The overlapping scales of bony fishes, e.g. Carp, are of different structure and consist of thin, flexible, translucent plates covered

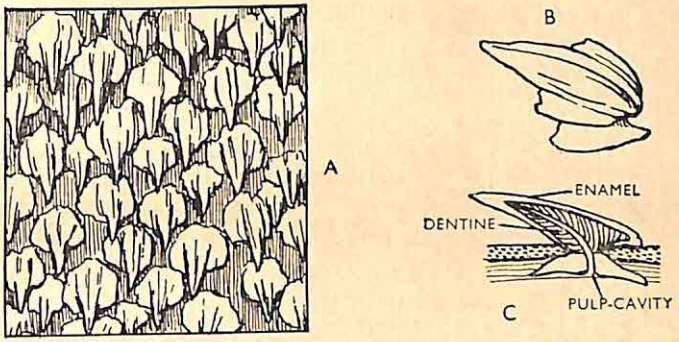


FIG. 74. Skin-scales of shark. A. Surface of skin as seen with hand-lens. B. Single tooth-like scale (magnified). C. Vertical section through scale (diagrammatic—after Borradaile).

by an outer layer of skin. They often show *growth-rings* that can be used to tell the age of the fish.

The horny scales of reptiles, e.g. turtles, and the scales on the legs of birds are different from those of either gristly or bony fishes.

In Book Three, Chapters II and III, we have already discussed *feathers* and *hair*, two other coverings that develop from the 'skin skeleton'.

TEETH

In Vertebrates the lining of the mouth is essentially an in-folding of the skin. In gristly fishes (e.g. sharks) the lining of the mouth is covered with the same spiny scales as the outer surface of the body, and in the region of the jaws these spiny scales grow very big and

become *teeth*. As one row of these teeth wears out, another row grows forward over the jaw to take the place of the old teeth. Such sharp, backward-pointing teeth are primarily for seizing and holding the food and not for chewing, and this is their main function in amphibians and reptiles. Teeth are never present in birds.

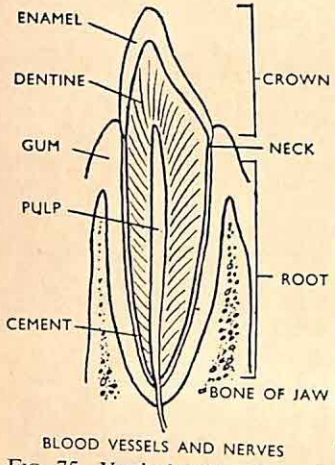


FIG. 75. Vertical section through human incisor tooth (diagrammatic).

One important characteristic of mammals is their *specialized teeth*. The individual teeth of all mammals are built on the same ground-plan, although they may differ widely in size and shape according to the animal's feeding habits. The part of the tooth that projects into the mouth is called the *crown*, while the part embedded in the jaw is called the *root*. The region level with the *gum*, where the crown joins the root, is called the *neck* of the tooth. The greater part of a tooth consists of *dentine* (or ivory†). Dentine, although it is very strong, is *elastic*, so that it can withstand sudden pressure without breaking. Dentine, however, is too soft to stand continuous wear, and the crown of the tooth is covered with a layer of very hard *enamel*.† Enamel is the hardest of all tissues, and in fully-grown teeth it consists almost entirely of mineral (or inorganic) matter. Although enamel is brittle, it is saved from breakage because it rests on the elastic dentine beneath. The enamel extends down to the neck of the tooth. The dentine of the root is covered with a thin layer of bone-like *cement* that helps to fix the tooth in its socket in the jaw. In the middle of the dentine is the *pulp-cavity*, containing a soft mass of blood-vessels and nerve-fibres, which enter through a tiny hole at the tip of the root (see Fig. 75).

Mammals have different kinds of teeth, specially adapted for different work (see Fig. 76). In *Man* there are four kinds of teeth:

(*a*) *Incisor teeth*—flattened, cutting teeth for biting. These are the four front teeth in each jaw. The incisors are well developed in gnawing mammals, e.g. rats, mice, cavies, squirrels. (*b*) *Canine*†

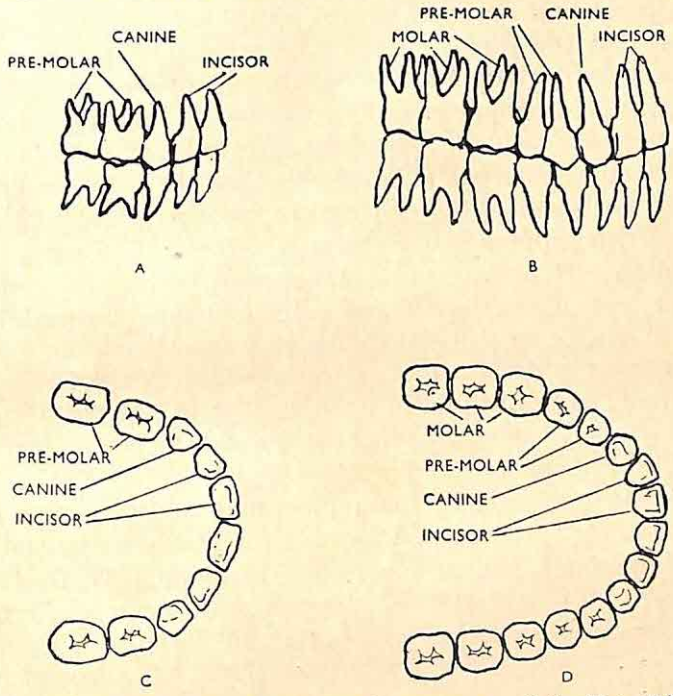


FIG. 76. Human teeth. A. Half-set of temporary teeth ('milk-teeth'). B. Half-set of permanent teeth ('second teeth'). C. Biting surface of temporary teeth. D. Biting surface of permanent teeth.

teeth—more pointed and stronger, two in each jaw, on either side of the incisors. These canine teeth are best developed in flesh-eating mammals (e.g. cats, dogs, tigers), being used for tearing up the food. (In Man the canines help the incisors in biting.) (*c*) *Pre-molar*† *teeth*—four small, grinding teeth in each jaw, placed in pairs behind each canine tooth. (*d*) *Molar teeth*†—six large back teeth per jaw for grinding up the food and mixing it with saliva: three on either side of each jaw. The molars and pre-molars

(or 'double' teeth) are well developed in herbivorous mammals, e.g. sheep, goats, cattle.

Man grows two sets of teeth—a *temporary set* during childhood and a *permanent set* when grown up. When a baby is about 7 months old the first front teeth appear. Other teeth appear gradually until by the time the child is about 2 years of age, the *temporary set* is complete, consisting of 4 incisors, 2 canines, and 4 pre-molars in each jaw. These 20 'milk teeth' serve the child until it is about 6 years old. During this time the jaws are growing longer, thus making room for the *permanent teeth* that are gradually developing inside the jaws. At about 6 years of age the child's first permanent molars appear, behind the temporary molars; and during the next few years the temporary teeth loosen and fall out, being replaced by the permanent teeth, until by the 15th year only the 3rd molars (or 'wisdom teeth') have still to appear. The complete set of permanent teeth, therefore, consists of 4 incisors, 2 canines, 4 pre-molars, and 6 molars in each jaw, a total of 32 teeth.

In order to summarize the number and kind of teeth a *dental formula* is used. The dental formula for Man's permanent teeth is: $2(\frac{2}{2}$ incisors $\frac{1}{1}$ canines $\frac{2}{2}$ premolars $\frac{3}{3}$ molars) Total = 32, or simply $2(\frac{2}{2}\frac{1}{1}\frac{2}{2}\frac{3}{3}) = 32$. Other examples are: human 'milk teeth', $2(\frac{2}{2}\frac{1}{1}\frac{2}{2}\frac{0}{0}) = 20$; dog, $2(\frac{3}{3}\frac{1}{1}\frac{4}{4}\frac{2}{3}) = 42$; sheep, $2(\frac{0}{3}\frac{0}{1}\frac{3}{3}\frac{3}{3}) = 32$; rabbit, $2(\frac{2}{1}\frac{0}{0}\frac{3}{2}\frac{3}{3}) = 28$; cat, $2(\frac{3}{3}\frac{1}{1}\frac{3}{2}\frac{1}{1}) = 30$; rat, $2(\frac{1}{1}\frac{0}{0}\frac{0}{0}\frac{3}{3}) = 16$; elephant, $2(\frac{1}{0}\frac{0}{0}\frac{0}{0}\frac{6}{6}) = 26$. It is interesting to note that the elephant's teeth are all molars (except for 2 incisors—the *tusks*†), but these molars are so big that there is only room in the jaw for two on each side at the same time. When one of these molars is lost another grows to take its place.

Examine any skulls that are available, and note the different types of teeth and their arrangement. Construct their dental formulae.

KINDS OF MOVEMENT IN VERTEBRATES

The chief ways in which vertebrate animals move are (*a*) by *swimming*, (*b*) by *walking*, and (*c*) by *flying*.

Fishes move by waves of muscle-contraction passing alternately along either side of the body, from head to tail (see Fig. 77). The unpaired fins act like the 'keel' of a boat and the paired fins are used in balancing and steering.

The snakes, which lack limbs, crawl by means of movements of broad scales on their under-surface. These scales are attached to

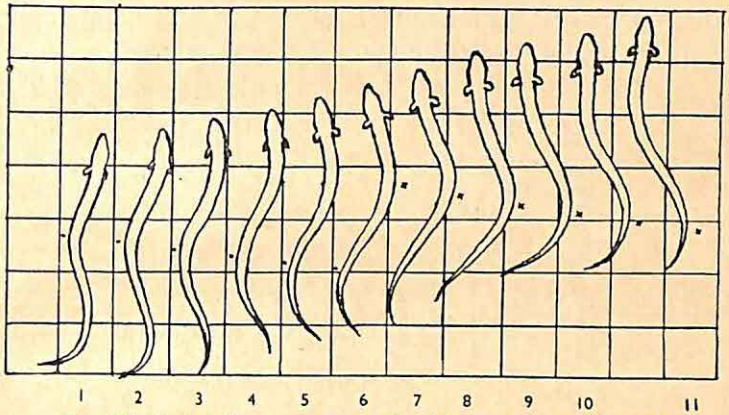


FIG. 77. Swimming-movements of a fish (one complete cycle).

the ribs, which are moved by muscles. The scales are raised in turn and thus push the snake along the ground.

In most vertebrate animals, *walking* involves the use of pairs of jointed walking-limbs. Most vertebrates walk on four limbs whose jointed bones act as levers for the pull of muscles, and two limbs support the weight of the body while the other two are lifted off the ground. Jumping animals, e.g. grasshoppers, frogs, and kangaroos, have exceptionally long and powerful hind legs.

In *flying* the wings make upward and downward movements, not unlike our own movements in swimming with the breast-stroke. Wing-movement during flight is, however, very complicated, and during the downward and forward stroke of the wings the feathers offer their maximum resistance to the air, and the bird lifts itself upwards. During the backward and upward stroke the wing-feathers offer less resistance to the air, which is allowed to flow

between them. Another type of flight is *gliding*, when the bird keeps its outstretched wings almost still and makes use of upward convection currents of air to sail about (with no loss of speed or vertical height) for long periods without beating its wings.

MUSCLES AND THEIR ACTION

In all animals, except the lowest, movement is brought about by *muscle-cells*, whose special function is *contraction*. These muscle-

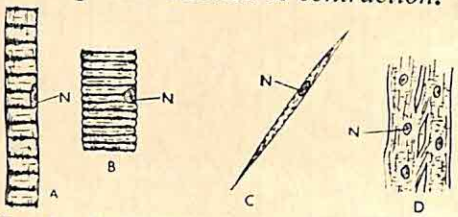


FIG. 78. Muscle-fibres (magnified). A. Part of voluntary (striped) muscle-fibre, relaxed (×130). B. Same voluntary muscle-fibre contracted (×130). C. Involuntary (smooth) muscle-fibre (×700). D. Heart muscle-fibres (×250). N. Nucleus.

cells are so long that they are also called *muscle-fibres*. Large groups of these muscle-cells make up muscular tissue or *muscles*.

Muscle-tissue is of three main kinds: (*a*) *voluntary muscle*, (*b*) *involuntary muscle*, and (*c*) *heart-muscle*.

Voluntary muscles are sometimes called *skeletal muscles* because they are mostly attached to bones and produce movements of the skeleton. Such muscles are usually controlled by the will, hence the name 'voluntary muscle'. Under the microscope, voluntary muscle is seen to consist of fibres that show a cross-striped appearance, hence another name for voluntary muscle is *striped muscle* (see Fig. 78). Every voluntary muscle consists of a soft, thickened, central portion (shaped like a cigar) with a tendon at either end attaching the muscle to two bones. Muscle-cells (which may be as long as 3 mm.) contract in length as the result of a *stimulus* received from the central nervous system through a motor-nerve. It is important to realize that when a muscle shortens its

length it becomes thicker, so that the *volume of the muscle remains the same*. When we say that 'a muscle contracts' we mean '*contracts in length*', so it is more correct to say that '*a muscle shortens and thickens*'.

The characteristic movements of voluntary muscle are *short, rapid contractions*, and this 'striped muscle' is always present in any part of an animal where rapid movement is required, e.g. in a mosquito's wing-muscles, which can contract and relax more than 500 times a second. Voluntary muscle, however, soon becomes *fatigued** (see also Book IIIA, pp. 108–109).

Involuntary muscle is sometimes called *smooth muscle* or *plain muscle* because it shows no cross-striped appearance under the microscope (see Fig. 78). Involuntary muscle is not attached to bones, but it is usually present in muscular sheets lining the internal organs, such as the food-canal and the blood-vessels, i.e. those parts of the body that are not under the control of the will. The characteristic movements of involuntary muscle are *slow, sustained contractions*; hence this 'smooth muscle' is useless for producing rapid movements. For example, the 'smooth', involuntary muscle of a frog's stomach may take 10 seconds to contract, as compared with 0·05 seconds for the contraction of the striped, voluntary leg-muscle of the same animal. Fatigue is slow to develop in involuntary muscle.

The *heart-muscle* of vertebrates differs from voluntary and involuntary muscle in its microscopic structure and in being able to sustain fairly rapid movements *without any nervous stimulus* (see Fig. 78). A frog's heart, for example, can go on beating for a long time after removal from the body. Heart muscle does not become fatigued.

Muscle-cells are, of course, living. When they contract they *do work*, and we know that this requires *energy*. The muscle-cells, therefore, are energy transformers, converting *chemical energy* into *mechanical energy*. The source of this energy is the glucose that is oxidized in the muscle-cells during the process of respiration. The actual chemical processes concerned have been found to be very

complicated indeed and there is no need for us to know anything of the details at this stage.

HOW LIVING THINGS BEHAVE

SENSITIVITY, IRRITABILITY, AND CO-ORDINATION

All living things respond to what is going on around them, i.e. they are *sensitive* to changes in their *environment* and can adjust themselves to these *external stimuli* by suitable *responses*. In order that a *stimulus* shall produce the *response* that will be of most benefit to the living thing, there must be *co-ordination*.† These are the three factors of *behaviour*: (*a*) *sensitivity—reception of a stimulus*; (*b*) *irritability—active response to the stimulus*, and (*c*) *co-ordination—adjustment of the response to the stimulus.*

'PLANT BEHAVIOUR'

Plants, like animals, respond to external stimuli, but as they do not move about to seek and capture their food or to escape from their enemies, these responses are quite slow. Such slow movements are usually carried out merely by *growth*, and there is no complicated arrangement of sense-organs, nervous system, jointed skeleton, and muscles such as we find in most animals. The chief external stimuli that influence plants are *gravity*, *light*, and *water*, and these are the main factors in 'plant behaviour'.

ANIMAL BEHAVIOUR

Unlike plants, most animals seek and capture their food and they are able to avoid their enemies. This usually involves much more sensitive arrangements for receiving external stimuli and more efficient methods to secure rapid response. We shall now trace the advance in sensitivity, response, and co-ordination in the animals we have studied in this Course.

In *Amoeba* and *Paramecium* there are neither sense-organs nor a nervous *system*. The whole surface of the cell is sensitive to external stimuli and the whole of its protoplasm responds. In

Hydra some of the cells are specialized to conduct impulses, and there is a network of nerve-cells that provides for co-ordination between the various body-processes, e.g. movement, food-capture, and digestion (see Fig. 79).

In *Hydra*, however, there are no actual *nerves*, nor any central controlling organ, merely a net-work of fine nerve-threads connecting the sensory-cells with 'muscle-tails' and nerve-cells.

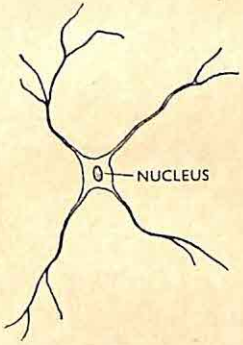


FIG. 79. Nerve-cell of hydra.

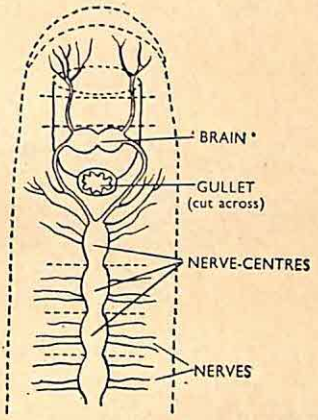


FIG. 80. Nervous system of earthworm (anterior portion only—after Dakin).

In the *earthworm* the nerve-cells are not scattered throughout the animal's body (as in *Hydra*) but they are grouped together to form a definite *nervous system*. There is a *ventral nerve-cord* consisting of small masses of nerve-cells (nerve-centres) joined together by nerve-fibres. These nerve-fibres are long, thin extensions of nerve-cells. Other nerve-fibres pass out from the nerve-centres to the various body-organs. The two largest nerve-centres are situated in the earthworm's head, forming a primitive 'brain' (see Fig. 80).

The nervous system of *insects* is built on the same plan as that of the earthworm—a ventral nerve-cord with a nerve-centre (or 'secondary brain') in each segment (see Fig. 44).

The nervous system of *Molluscs*, e.g. the snail, is more concentrated, there being three large nerve-centres controlling the head, the digestive organs, and the 'foot' (see Fig. 81).

THE NERVOUS SYSTEM OF VERTEBRATES

The nervous system of vertebrate animals is quite different from that of invertebrate animals. In all Vertebrates the *central nervous system* consists essentially of a hollow tube—the *spinal cord*—enlarged at its anterior end to form the *brain*. Unlike the *ventral* nerve-cord of Invertebrates, this central nervous system is *dorsal* to the food-canal and it is protected by the backbone and the skull. We have already discussed in outline the structure and function of the vertebrate brain (Book Three, Chap. II).

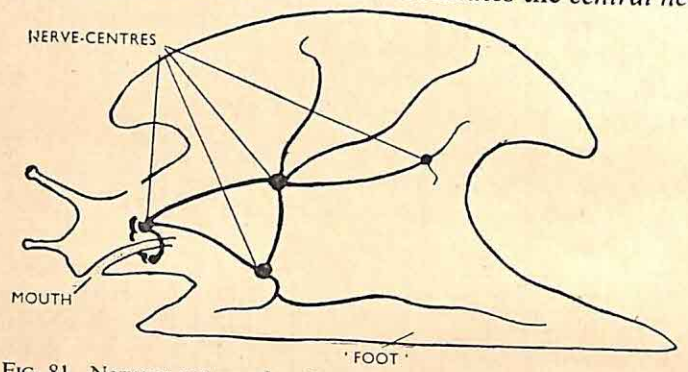


FIG. 81. Nervous system of mollusc (diagrammatic—after Brightwell).

The *central canal*, which runs from end to end of the spinal cord, opens out into several cavities inside the brain, and both central canal and brain-cavities are filled with liquid. As a young vertebrate animal develops from the fertilized egg-cell, the spinal cord is the first part of the central nervous system to develop, and then the anterior end of the spinal cord begins to swell out to form the brain. Three such swellings are formed in all young Vertebrates —the *fore brain*, the *mid brain*, and the *hind brain*. From the primitive *fore brain* develop the two halves (the *cerebral hemispheres*) of the *big brain* (or *cerebrum*). The more highly developed the big

brain, the more intelligent the animal, since this part of the brain is the seat of all voluntary action and conscious sensation. The most characteristic feature of the human brain is its exceptionally large and much-folded *cerebral hemispheres*. The *mid brain* controls sight and hearing, and it is also concerned with the control

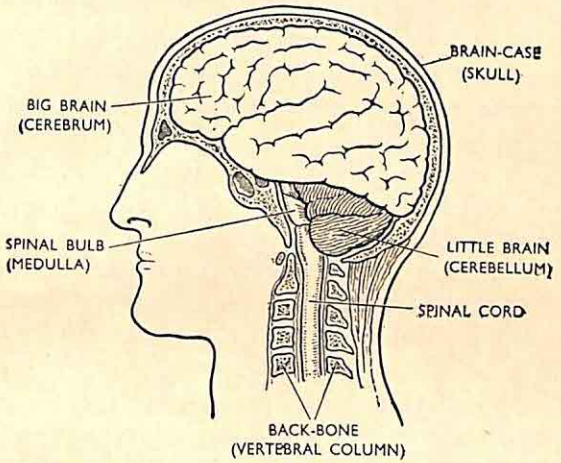


FIG. 82. Section through skull and part of backbone (after Cullis and Bond).

of posture,* the way in which the body is carried. The *hind brain* develops into the *spinal bulb* (medulla†) and the *little brain* (cerebellum). The *spinal bulb* controls the unconscious movements of breathing, movements of the digestive organs, heart-beat, and the activity of certain glands. The *little brain* appears to co-ordinate the complex actions of all the muscles concerned with movement, e.g. a mammal with a damaged cerebellum cannot stand. The little brain receives messages from the organ of balance (which is part of the inner ear) and from the small internal sense-organs in the muscles and joints.

In addition to this *central nervous system*, Vertebrates have also a *sympathetic nervous system*† consisting of a double chain of nerves running parallel to and on either side of the spinal cord. The branch nerves from this sympathetic nervous system run to the

involuntary muscles and the glands that are not under conscious control, e.g. the glands and blood-vessels of the food-canal, the iris-muscle of the eyeball, and the blood-vessels in the skin (see Book Three, Fig. 75).

THE SENSE-ORGANS—RECEPTORS†

An animal's behaviour depends mainly upon what is happening in the outside world around it. For example, when an animal sees food it makes the movements necessary to capture it, or if it hears a frightening noise from one direction, it moves away in another direction. A vertebrate animal gets information about what is going on in the outside world by way of the *eyes*, the *ears*, the *nose*, and by way of sensitive cells in the *skin*. Sensory nerves carry impulses from these *receptors* to the central nervous system. In this Course we shall only have time to discuss the sense-organs of Vertebrates. taking Man as our main example.

SIGHT—THE EYE

The *eyeball* rests on a cushion of fat within its bony *eye-socket*, and it can be turned up and down, and from side to side, by six flat strips of muscle. Skull-bones project above and below, while the nose also gives some further protection. The *eye-lids* provide extra protection, closing by *reflex action* when the eyeball is touched or when an object moves near the eye. The front of the eyeball is transparent and is kept wet by a strongly *antiseptic* liquid (the *tears*) produced by a small *tear-gland* in the outer corner of each eye-socket. The transparent window in front of the eyeball (the *cornea*†) is washed free from bacteria-laden* dust particles each time the eyelids are closed and opened. The tears drain away into the nose-cavity through a tiny duct that opens from the inner corner of each eyelid. (In crying, uncontrolled laughing, and violent coughing, the tear-gland produces more liquid than this tiny duct can drain away into the nose, and tears roll down the cheeks.)

The eyeball is enclosed by a tough, fibrous outer wall, lined with

a black, non-reflecting layer except in front, where the outer layer is thin and transparent.

Although the human eye resembles a photographic camera (see

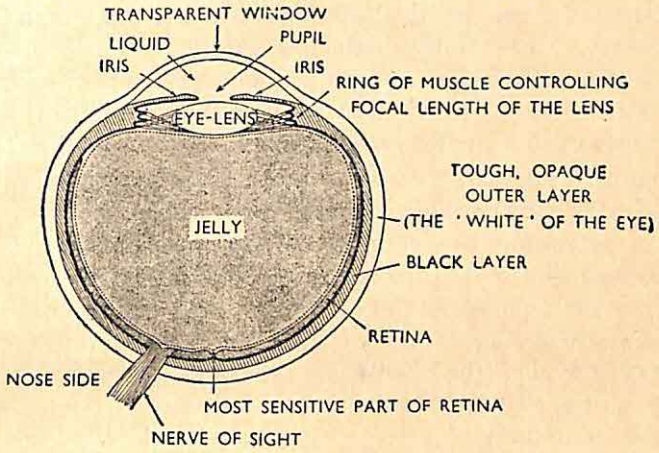


FIG. 83. Right eye in horizontal section (diagrammatic—after Cullis and Bond).

Book Three, Chap. IV), it is more sensitive, more adaptable, and more complex than the best cameras. Rays of light enter the eyeball through its curved transparent window (the *cornea*) and

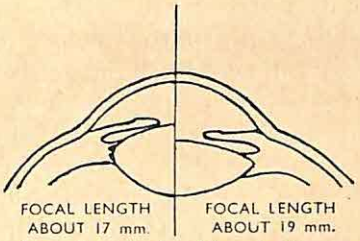


FIG. 84. How the eye-lens alters its focal length (diagrammatic—after Roaf).

are *refracted* to form an image on the *retina*. These light-rays are bent more by the surface of the cornea than by the eye-lens, whose main function is to *change* the focusing power of the eye by adjusting its own *focal length* (see Fig. 84). The *retina* is about 0·02 in. thick and, except in the shallow pit (*fovea*†) in the

middle of the *yellow spot*† (*macula lutea*†), nerve-fibres and nerve-cells interfere with the sharpness of the image focused on the retina. Although this would be a serious drawback in a camera, it is a great practical advantage in the eye, for if everything we saw was equally distinct it would be very difficult to concentrate our attention on one particular object. So when we 'look at' an object we turn our eyeballs so that an image is formed in each eye on this small *area of sharpest vision*. In the retina a delicate purple pigment (*visual purple*†) is bleached when light falls on it, thus giving a light-picture. This photochemical effect generates nerve-impulses running to the brain, where these impulses are translated into the sensation of sight.

The *iris*, a circular sheet of coloured muscle, corresponds to the '*stop*' of a camera, since it can expand or contract and thus change the size of the *pupil* (the opening in the centre of the iris). In this way the amount of light entering the eye can be controlled. If you stand in a shady place with your back to the light, and look at your eye in a hand-mirror, you will see that the pupil is large, so as to admit more light. If you then turn to face the light, you will see the iris contract. This reflex reaction protects the delicate retina and its sensitive nerve-cells from the shock of a sudden increase in the amount of light that enters.

The chamber of the eye in front of the eye-lens is filled with a watery liquid (*aqueous humour*†). Immediately behind the iris is the double-convex *eye-lens*, a transparent, elastic body. In mammals the curvature of this lens can be changed by the movements of rings of muscle at the circumference of the lens, so that the lens can focus light rays from near objects to produce a clear image on the retina (see Fig. 84). The large back chamber between lens and retina is filled with a transparent, jelly-like substance (*vitreous humour*†).

DISSECTION OF EYE

Examine the eye of an ox, sheep, or goat from the abattoir. Notice that, except in front, the eyeball is surrounded by fat,

which forms a protective cushion in the bony eye-socket. Carefully remove as much of this fat as possible and identify (*a*) the six *eye-muscles*, (*b*) the *nerve of sight* (*optic nerve*).

With scissors, or a sharp knife, separate the front half of the eye from the back half and identify (i) the front chamber, in front of the eye-lens, filled with a watery liquid, and (ii) the back chamber, behind the eye-lens, filled with a jelly-like substance (you may be able to remove this, together with the eye-lens), (iii) the *eye-lens*, (iv) the *iris*, (v) the *pupil*, (vi) the tough, white, outer *layer* (*sclerotic*†), (vii) the thin *black lining* (*choroid*†), (viii) the thin grey *retina* (which can be peeled off with care).

Examine an eye that has been hardened in 40 per cent. formaldehyde overnight. Cut it in two, from front to back, in vertical section. Make a diagram of what you see. Also examine an eye-lens that has been removed undamaged.

HEARING—THE EAR

In Man, as in most other mammals, the visible *external ear* is a gristly structure that collects sound-waves and directs them down a tube on to the *ear-drum*. Some mammals, e.g. the horse and rabbit, can actually turn the external ear so as to collect sounds coming from different directions. (The horse, which moves its external ears a lot, has 17 muscles for moving each ear; Man has only 9 such muscles and they are all useless, except in rare cases.)

A narrow, slightly curved tube (about one inch long in Man) leads inward from the external ear to the *ear-drum*. This tube is lined with outward-projecting hairs and wax-producing glands that help to protect the ear-drum from insects and harmful solid matter. The ear-drum consists of a very thin membrane (only a few thousandths of an inch thick in Man) stretched across the inner end of the tube leading from the external ear. Being so very thin, the ear-drum is easily damaged.

The rest of the ear—the *internal ear*—is entirely enclosed in the bone of the *ear-cases*, which form part of the skull. The *internal*

ear is divided into two parts: (*a*) the *middle ear*, and (*b*) the *inner ear* (see Fig. 85).

The *middle ear* is a cavity in the bone of the skull, filled with air, and connected to the back of the throat by a narrow passage (the

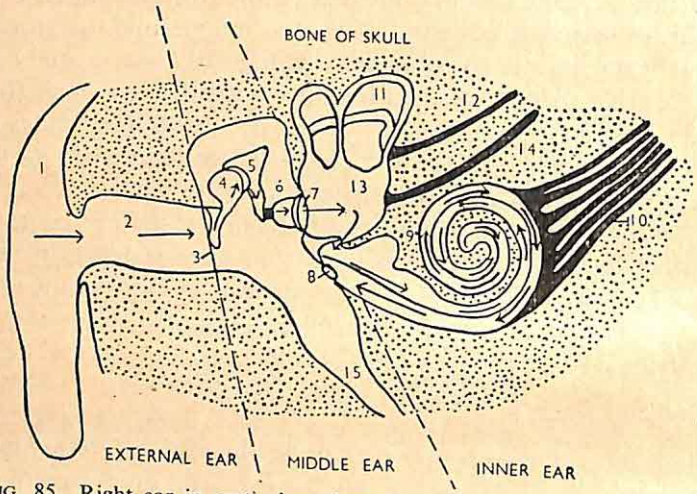


FIG. 85. Right ear in vertical section (diagrammatic). 1. External ear. 2. Tube leading to ear-drum. 3. Ear-drum. 4, 5, 6. Chain of bones. 7. Oval window. 8. Round window. 9. Organ of hearing. 10. Nerves of hearing. 11. Semicircular canals. 12. Nerve of sense of direction of movement. 13. Organ of balance. 14. Nerve of sense of balance and position. 15. Eustachian tube leading to throat.

Eustachian tube†). By means of this tube the air pressure on the inner surface of the ear-drum is kept the same as the air-pressure on its outer surface. (You can now understand why, in war-time, people are advised to keep their mouths open when bombs are falling nearby. This enables the sudden increase of air-pressure to act on both sides of the ear-drum, thus saving this thin membrane from being burst inwards.)

A *chain of three small bones*, hinged together, stretches across the middle ear and carries vibrations from the ear-drum to the *inner ear*, which contains the true *organ of hearing*, enclosed in the bony mass of the skull. The external ear and the middle ear,

therefore, are only concerned with receiving sound-waves and passing them on to the inner ear. (The *inner ear*, however, contains some other organs besides the organ of hearing: one part of

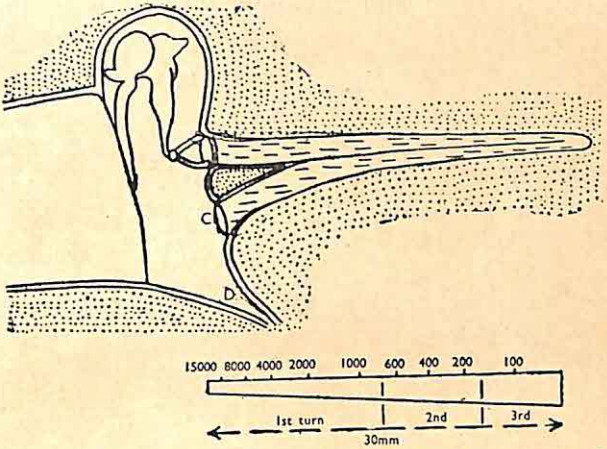


FIG. 86. Diagrammatic section through ear with spiral tube of cochlea unrolled (after W. H. White). The lower diagram is a plan of the elastic membrane that divides this tube into two passages marked with the frequencies of the vibrations to which the different parts of this membrane respond.

it is the organ of balance and position, which we shall discuss later.) There are two small openings, closed by membranes, between the middle ear and the inner ear. They are called the *oval window* and the *round window* (see Figs. 85 and 86).

The actual *organ of hearing* consists of a spirally-coiled passage shaped like a tiny snail-shell and filled with a liquid; it is called the *cochlea*.† The vibrations of the ear-drum are passed on and magnified by the chain of bones across the middle ear to the membrane of the *oval window*, and similar vibrations are set up in the liquid filling the *cochlea*. These vibrations travel through the liquid up to the top of the spiral tube and then down again to the *round window*. (Hence, liquids being incompressible, as the oval window moves inwards the round window moves outwards.)

Embedded in the elastic membrane separating the two halves

of the spiral tube (see Fig. 86) are about 24,000 tiny fibres varying in length and tension. The largest of these is only $\frac{1}{15}$ in. long. This part of the ear has been compared to a tiny but very complicated piano (without the keys), and it is supposed that, when the liquid

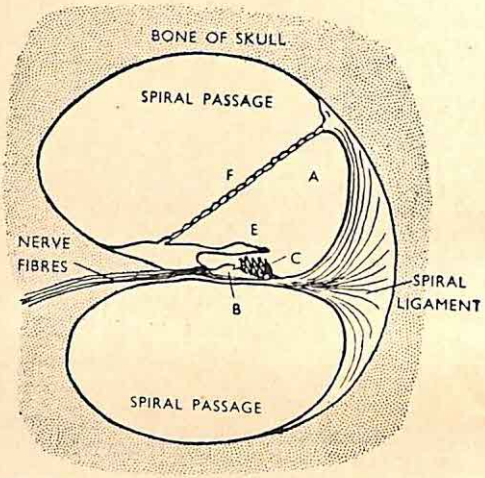


FIG. 87. Diagrammatic section across organ of hearing (magnified). This shows an enlarged cross-section of the spiral passages shown in Figs. 85 and 86. A. Third, narrower duct lying between the two spiral passages in the cochlea. B. Elastic membrane supporting rods with sensitive hair-cells (C). E. Membrane forming 'shelf' overhanging the sensitive hair-cells. F. Membrane separating duct A from spiral passage.

in the spiral tube vibrates, those fibres 'in tune' will vibrate, too, in sympathy (i.e. they will *resonate*) and transmit tiny electrical impulses to the brain. The brain recognizes these impulses as notes of varying pitch according to the particular nerve-fibres that carried them. 'Within the small volume of the *cochlea* lie 24,000 minute electrical machines and from it emerges a cable of 24,000 conducting nerves. This machinery is so complicated that science has not yet completely discovered its workings.'[1]

[1] From Jeans's *Science and Music*.

EXPERIMENTS ON HEARING

(i) Listen to the note of a whistle whose pitch can be gradually raised. Notice that when the pitch of the note gets very high it becomes more difficult to detect and finally it cannot be heard at all. Notice that some individuals can hear 'higher' notes than others (e.g. some people cannot hear the very high-pitched squeak of a flying bat).

(ii) Place your flat hands, with their palms facing backwards, on either side of the head, just in front of the ears. Stand in front of a loudly-ticking clock and close your eyes. Where does the sound appear to come from? Your hands 'shade' the ear-passages from sound coming from the front; hence the sound waves can only travel round the back and sides. Use your hands as extensions of your external ears and notice the increased loudness of sounds coming from in front of you.

SENSE OF BALANCE AND POSITION

The inner ear contains the three *semicircular canals*†, which are concerned with the sense of position and balance of the body. Each of these three semicircular canals lies in a plane at right angles to the other two, and at one end of each canal there is a swelling that contains nerve-cells (see Fig. 88). The semicircular canals are filled with liquid, and when the head is moved in any direction this liquid drags behind a little and affects the nerve-cells, which then send an impulse to the brain. In the brain this impulse gives rise to the sensation of *movement* in a particular direction.

It is clear that sensations of movement depend upon the *movement* of the liquid in the semicircular canals. But we can also sense the *posture* of the head even when this is not moving. This power is given by another part of the organ of balance that lies below the canals. This part of the inner ear consists of two connected sacs containing tiny chalky particles attached to sensitive cells. These solid particles, when pulled downward by gravity,

pull on some sensitive cells more than on others. In the brain the impulses coming from these cells give rise to sensations of *position*. The organ of balance is helped in its work by other sense-organs. Thus our eyes tell us a great deal about our movement or position in relation to surrounding objects.

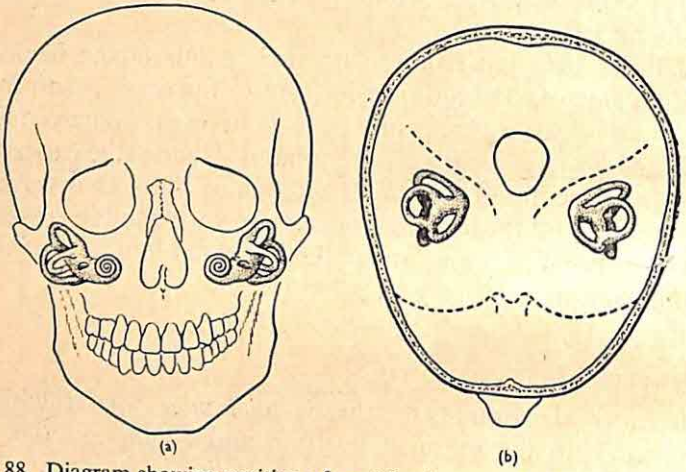


FIG. 88. Diagram showing position of semicircular canals in head. (*a*) Viewed from front of head. (*b*) Viewed from above head (after Roaf).

INTERNAL SENSE-ORGANS

There are other sense-organs, scattered throughout the body, that report to the brain what is going on in the voluntary muscles, tendons, joints, and other parts of the body. For example, these internal sense-organs tell you when you are tired, when you are hungry or thirsty, or when your stomach is over-full. Similarly, it is not necessary to have your eyes open to know the movements and position of your limbs because this information can be supplied by internal sense-organs.

SENSES OF TASTE AND SMELL

The *organ of taste* is the *tongue* and the *organ of smell* is in the upper part of the *nose*. These two organs nearly always work

together. In fact, it is usually difficult to decide whether a particular sensation is a taste or a smell. When one has a 'cold in the nose' one cannot smell, and at the same time food has very little

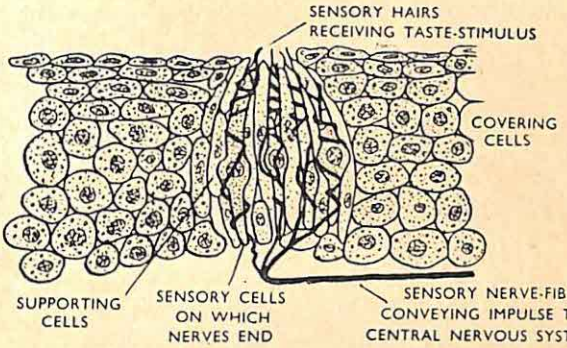


FIG. 89. Section through taste-bud in human tongue (magnified--after Hatfield).

taste. This is because most of our ideas of the 'taste' of foods also include their characteristic smell.

We taste with special groups of sensitive nerve-cells called *taste-buds*,† situated in tiny projections on the upper surface of the

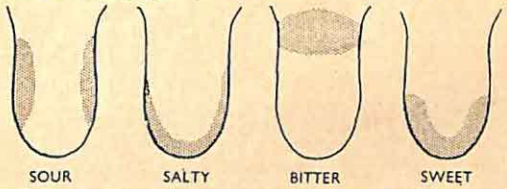


FIG. 90. Diagram showing taste-sensitive areas of human tongue.

tongue and on other parts of the mouth-cavity. These taste-buds are connected with the brain by nerve-fibres (see Fig. 89).

These *taste-buds*, which are only sensitive to substances that have *dissolved* in the saliva, are of four kinds, each corresponding to one of the four pure tastes—*sweet*, *sour*, *bitter*, and *salty*. Most everyday tastes are combinations of these fundamental tastes. The four kinds of taste-buds are not distributed evenly over the tongue (see Fig. 90). Thus we taste sweet things best with the tip of the tongue, where the 'sweet taste-buds' are most numerous,

while bitter tastes affect mainly the back of the tongue. The sour and the salty taste-buds are mainly in the sides of the tongue, the sour taste-buds being most numerous about the middle of each side and the salty taste-buds being towards the tip of the

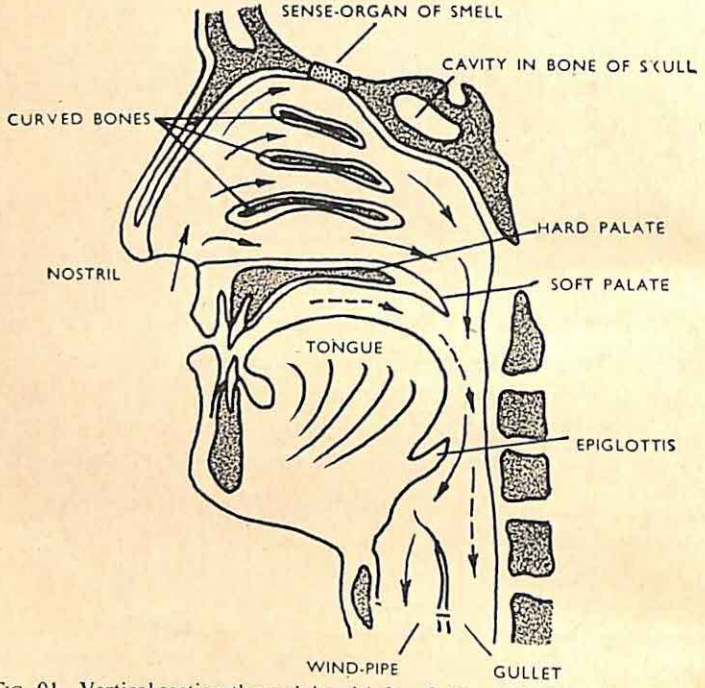


FIG. 91. Vertical section through head (after Cullis and Bond). Dotted arrows show path of food and drink. Black arrows show path of inspired air.

tongue. We recognize the impulse received in the brain from the different taste-buds as a particular kind of taste.

We *smell* by means of special sensitive hair-cells situated on a square *smelling-patch* in the upper part of the nose-cavity (see Fig. 91). These cells are connected to the brain by nerve-fibres. As air passes through the nose, these sensitive cells are stimulated by any impurities that have a definite smell, and a stimulus is sent to the brain along the nerve-fibres. We recognize this stimulus as a

particular smell. (Your teacher will tell you how to do some simple experiments on taste and smell.)

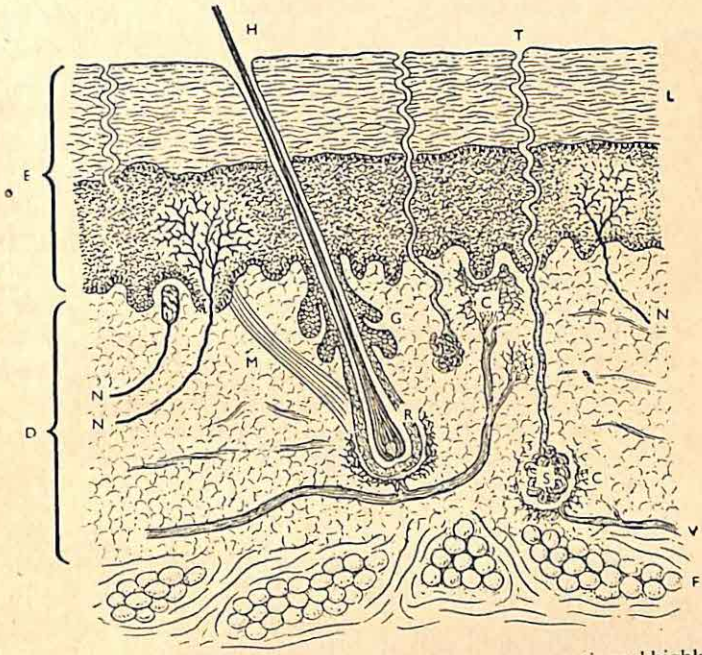


FIG. 92. Vertical section through human skin (diagrammatic and highly magnified). C. Blood capillaries. D. Inner layer (dermis). E. Outer layer (epidermis). F. Fat. G. Oil gland. H. Hair. L. Layer of flat, horny, non-living cells. M. Hair-muscle. N. Nerve fibres and sensitive nerve-endings. R. Hair-pit. S. Sweat-gland. T. Sweat-pore (opening from sweat-duct). V. Larger blood-vessel.

SENSES OF FEELING, TOUCH, TEMPERATURE, AND PAIN

We get much information about our surroundings through the sense-organs of *touch* situated just beneath the outer *skin* over most parts of the body-surface (see Fig. 92). When we touch anything, even with our eyes closed, we can say whether it is hot or cold, hard or soft, and whether it hurts or not.

This is because the skin has four kinds of sense-organs, each

sensitive to one only of the four types of sensation: (i) *warmth*, (ii) *cold*, (iii) *pain*, and (iv) *touch*. Nerve-fibres carry stimuli from these sense-organs to the brain and to the spinal cord. (Your

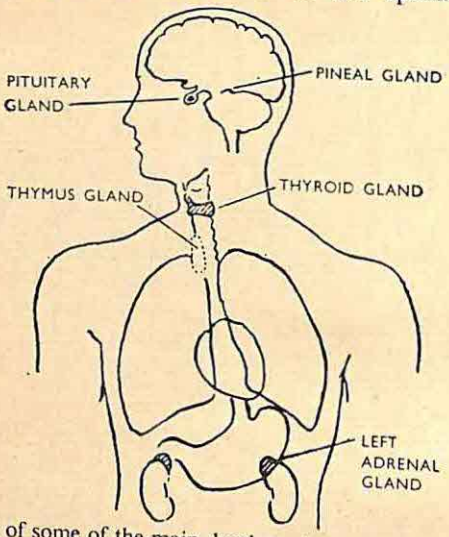


FIG. 93. Position of some of the main ductless glands (after Holmes and Gibbs).

teacher will tell you how to do some simple experiments on the sense-organs of the skin.)

The sensation of *pain* helps to protect us against injury, e.g. a limb is bent and drawn away from the cause of the pain. Pain also tells us when something is wrong with the body, and slight damage can be dealt with before more serious harm results, e.g. pain tells us that a tooth is decaying and we can have it attended to before it gets worse. Pain also informs us just *where* the trouble is. Pain often makes us *rest* the injured part and so helps healing and repair (for rest is one of the best aids to healing).

CHEMICAL CONTROL—HORMONES

Internal co-ordination of different parts of the body is not always carried out by the nervous system. Co-ordination is also carried out by the circulation of chemical substances with the

blood. These '*chemical messengers*' are called *hormones* and they are passed into the blood-stream by special *ductless glands*† (see Fig. 93). For example, the pancreas does not pour out pancreatic juice continuously, but only when the small intestine contains food to be digested. Whenever food enters the small intestine from the stomach, cells in the wall of the small intestine produce a hormone that is passed into the blood-stream and carried round the body. When blood containing this hormone reaches the pancreas, it supplies a stimulus to that gland, which pours out pancreatic juice. The same hormone may also stimulate the flow of bile from the liver.

In female mammals, when a fertilized egg-cell begins to develop in the body, another hormone is produced that stimulates the growth of the milk-glands ready to supply milk for the young one after birth.

Another very important hormone is *adrenalin*,† produced by the *adrenal bodies*,† a pair of ductless glands lying just anterior to the kidneys. Under great excitement, e.g. anger or fear, the adrenal glands pass adrenalin into the blood-stream, and as the hormone is distributed to every part of the body it prepares almost every organ for special efforts. For example, the heart-beat becomes more rapid, the capillaries become wider, the amount of glucose in the blood increases, the sweat-glands become more active, and the activities of the food-canal are slowed down.

The *thyroid gland*,† in the neck, produces a hormone that regulates growth and development. A deficiency of this hormone slows down the heart-beat, blood-circulation, and oxidation of glucose in the body-tissues during respiration. Thyroid deficiency in young people restricts growth and bone development. Nervous activity is also reduced, so that mental development becomes backward.

The *pituitary gland*,† situated beneath the brain, produces another hormone that also controls growth. Most human 'giants' have unusually active pituitary glands, while a lack of this hormone produces 'dwarfs'.*

You will notice that for rapid action the body uses nervous control and that when the action is slow and prolonged it is controlled chemically by hormones.

Hormones are rather like vitamins in that exceedingly small amounts produce very striking results. In fact, it has been suggested that 'vitamins may be regarded as hormones that the body cannot make for itself'.

ANOTHER TYPE OF CHEMICAL CONTROL

A different type of chemical control is carried out in animals by the waste products, carbon dioxide and urea, but these substances are not called hormones. For example, the diameter of the blood-capillaries varies according to the amount of carbon dioxide in the blood. When the body-muscles are resting, there is little carbon dioxide in the blood and the capillaries are very narrow. When the muscles are working, the blood receives more carbon dioxide and the capillaries widen, thus increasing the blood supply. At the same time, when this blood—containing more carbon dioxide—reaches the brain, nervous impulses are sent to the muscles between the ribs and to the diaphragm, and the breathing movements become deeper and more rapid.

Similarly, an increased amount of urea in the blood-stream stimulates the kidneys to greater activity in order to remove this excess of urea from the blood.

INSTINCT AND INTELLIGENCE

The behaviour of the lower animals is guided mainly by *instinct*; that is to say, their ways of behaving in various situations are *inherited*; they do not have to *learn* what to do. A spider does not have to learn how to spin its web, or a bird how to build its nest. In so far as it is inherited, and not learnt, instinctive behaviour is similar to reflex response, but it differs from the latter in some important respects. When a dog jumps on hearing a sudden loud noise, this is a reflex response to the stimulus of the sound. But if the dog hears the squeak of a rat its response is not

merely to the stimulus of the sound, but also to what the sound *means* to the dog. It is a signal meaning the presence of a rat, and the dog's activity is much more than the mere contraction of a group of muscles; it is directed towards finding and catching the rat. There is every reason to believe that, in the higher animals at any rate, instinctive behaviour is accompanied by consciousness, and that quite complicated mental processes may be involved. But the essential feature is that the lower animal knows what to do without having had to *learn* it.

The behaviour of the lower animals is mainly *instinctive.* When an animal is able to *learn by experience*, it is said to show *intelligence.* Experiments show that even such lowly animals as insects are not completely without intelligence. Bees, for instance, can learn that honey is to be obtained from yellow containers but not from those of another colour. In all animals below mammals intelligence probably plays a quite minor part in their activities. Dogs and cats, however, quickly learn a great many simple things, such as to come and be fed when called. Experiments have shown that they can learn such things as how to release themselves from a cage by pulling a string that passes over a roller and is attached to the catch of the door.

Man is distinguished from the lower animals not only by his greater intelligence, but by his power of *reasoning.* The difference between intelligence and reason can be illustrated in this way. A monkey learns to open the door of a cage by pulling a string. It has discovered, in the first place by accident, that the door opens when the string is pulled. Being *intelligent*, it remembers this next time, and instead of trying to force its way out of the cage, looks for the string and pulls it. If the monkey were capable of *reasoning*, it would not only discover that pulling the string opens the door, but would understand *why* it does so. In other words, it would be able to form general ideas about mechanical devices and how they work. It is clear that an animal capable of *reasoning* can behave much more successfully in unaccustomed situations than can an animal that is merely intelligent. Instead of blundering* on the

chance of finding the right solution, it can think out how to modify the actions that it has found successful in different but somewhat similar situations, so as to suit the new situation. Experiments have demonstrated some power of reasoning in apes, and it appears to exist to a very slight extent in cats and dogs.

Man, however, clearly stands by himself among all other animals in the extent of his powers of reasoning. This has been greatly helped by his capacity for *speech*, for words enable us to formulate* much more complicated ideas than we could without their use, just as a long multiplication sum would be impossible without symbols to express the different numbers. Moreover, *language* enables people to share their ideas and experiences and hand them down from one generation to another. This is one of the most important factors in *education*.

(See *General Science Workbook*, Book Two, pp. 81–82, for learning exercises on this chapter.)

Chapter IX

GROWTH AND REPRODUCTION IN LIVING THINGS

All living things grow, and in this chapter we shall discuss and illustrate the scientific meaning of this word *growth*. Its simplest, everyday meaning is 'increase in size'. Many non-living things can grow in this sense; thus a bar of metal 'grows' in length when it is heated; a crystal of common salt 'grows' in size when it is placed in a saturated solution of salt which is then concentrated by evaporation.

When a crystal grows, it merely increases in size and weight. It keeps the same geometrical shape and its increase in size is caused by the addition of new molecules to its surface. If these molecules are removed again (by *solution*), the crystal decreases in size. Again, a heated bar of metal 'grows' because the molecules in it get farther apart. When it is cooled it returns again to its original shape and size. When, however, a living thing grows, there is an increase in size and (usually) of weight, but many other changes take place also. For example, *new (and often different) kinds of cells are formed*; the organism changes in *form* both internally and externally, new material is taken into the body and there *assimilated* into complex new substances. True growth in living things is the making of new protoplasm and new cells, together with a change in the structure and function (or *differentiation*†) of those cells. It is a *non-reversible* process.

Although growth in animals and plants is much the same process, there is one important difference. Plants grow in *length* mainly at their ends—at the tips of roots and in the terminal buds at the tips of shoots. There may be also, in some trees and shrubs, an increase in *thickness* of stems and roots, due to a special kind of *secondary* growth; this, however, is also localized. Animals do

not grow in this way; they show an 'over-all' growth, all the organs growing at the same time (though at different rates) until the body is adult, or *mature*. Thus a man 'stops growing' at the age of about 20 years, while a tree 1,000 years old may still be showing active growth at its stem and root-tips and in the thickening of its trunk. The difference is not always so marked as this; in many *annual* plants* the growing stem-tips develop into flowers, and the plant as a whole may be quite short-lived.

It is not strictly true, either, that an animal 'stops growing' at maturity. It is true that it may then get no larger in size or become different in form, but some obvious growth is still going on, e.g. in the surface hair and nails or claws, and it is constantly making new blood-cells and new skin-cells. It is probably true that every part of the adult body is constantly wearing out and being replaced by newly-grown tissue. Thus in the adult animal and in the mature parts of plants it is better to say that growth is still going on, but at a rate only sufficient to replace normal wastage.

THE GROWTH OF PLANTS

We can easily study growth by doing experiments with seeds. Every seed contains a tiny *embryo* plant, together with a store of organic food. Given suitable conditions, the embryo rapidly grows into a seedling and finally into the mature (but still growing) plant. The early growth of a seed is called '*germination*' and we studied this process in Book One, Chapter IV.

GROWTH IN DRY-WEIGHT

We have already learnt that true growth is accompanied by an increase in cell numbers and by the making of new protoplasm. We should expect, therefore, an increase in 'dry-weight' (i.e. the weight of a plant after water has been driven off). We should also expect to find an increase in *fresh-weight*, but, especially in plants, this changes rapidly apart from growth (e.g. during wilting). *Changes in the dry-weight of plants provide the best measure of growth.*

Demonstration. Take 100 rice grains[1] (padi) and heat them in an oven for 24 hours at 100° C.; cool in a desiccator, and then weigh them. This gives you the dry-weight of an average sample of 100 rice grains. Now take four similar samples of 100 rice grains and determine their dry-weights in the same way *after germination in the light* for (*a*) 3 days, (*b*) 7 days, (*c*) 14 days, (*d*) 21 days. The seeds should be soaked in water for one day and then set to germinate on wet filter-paper in deep glass jars with loose-fitting lids. Keep them moist during the growth period and afterwards dry them at 100° C. for 24 hours while still on the filter-paper (to which the roots will stick). The dry-weight of an average sample of the same filter-paper must be subtracted from each result.

In the early stages of germination, you will find that the seedlings (although growing) *lose* in dry-weight *as a whole*. This is because respiration is going on, causing the oxidation of some carbohydrate while food is being transferred to the shoots and roots from the reserve food-material in the seed. (Actually the seed loses in dry-weight while the shoots and roots gain.) Later (when the leaves have become green) the *whole seedling* begins to gain in dry-weight, and as growth proceeds this gain becomes progressively more rapid. This is due to the manufacture of new food by the green leaves, resulting in the growth of new cells.

GROWTH IN LENGTH

When a seedling grows it increases in height or length as well as in dry-weight. For some purposes it is useful to measure growth as an increase in length, but it is important to realize that growth in dry-weight may not always be accompanied by an increase in length; similarly, growth in length may take place without any change in dry-weight. This is because the size of an organism is due not merely to the *number* of cells in it, but also to the *size* of these individual cells, and this size may be greatly altered by changes of water-content.

Demonstration. Take three samples, each of 20 cow-pea seeds.

[1] Wheat, barley, oats or millet may be substituted for rice.

Dry one sample in an oven at 100° C., cool in a desiccator, and then weigh. This gives the initial* dry-weight of 20 seeds. Germinate the others, keeping one set in the dark and one set in the light. After one or two weeks determine the dry-weight of each set.

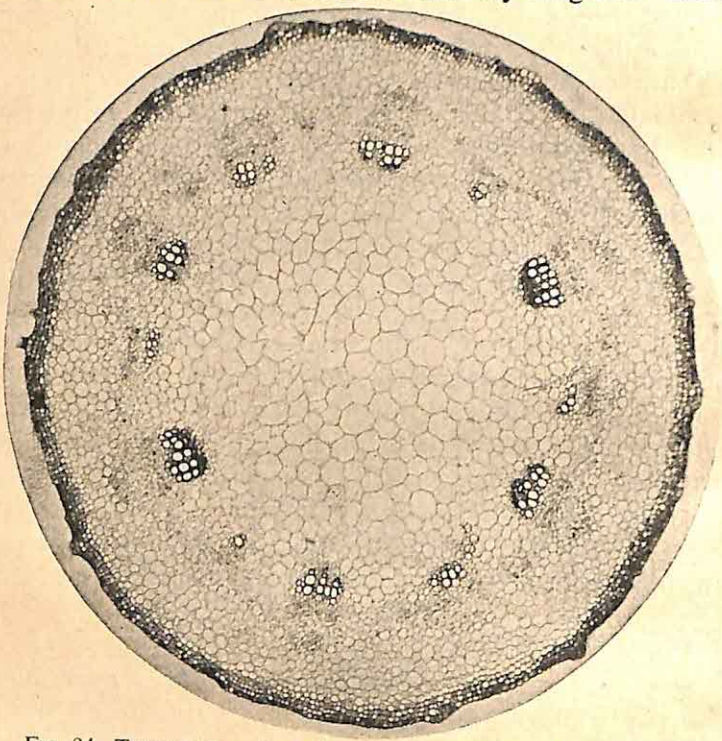


FIG. 94. Transverse section of young sunflower stem (a dicotyledon), showing vascular bundles arranged in a ring (highly magnified).

Notice that the set in the light has grown in length and has also increased in dry-weight. The set in the dark has *grown very much more in length* but has actually *decreased in dry-weight*. Explain these observations.

GROWTH IN THICKNESS (IN PLANTS)

In Book One, Chapter IV, we showed that *all* the growth in *length*

of a root takes place close behind the root-tip and that the growth in *length* of a stem is spread over a considerable length of the young stem. We shall now consider how the stems and roots of woody dicotyledons grow in *thickness* after the first growing season.

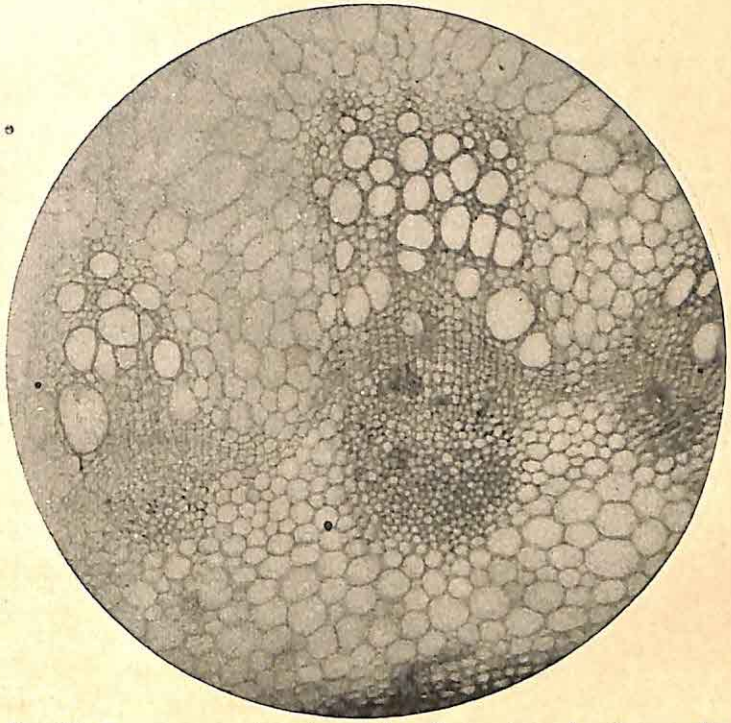


FIG. 95. Transverse section of part of sunflower stem, showing vascular bundles and cambium layer (more highly magnified).

A cylindrical layer of delicate cells, *cambium*,† is formed between the wood (*xylem*) and the bast (*phloem*). These cambium cells divide rapidly, mostly by walls tangential* to the central cylinder. New xylem cells are formed on the *inside* of the cambium layer and new phloem cells on the *outside*. (The cambium, however, always produces more xylem than phloem.) In this way the cambium provides more wood-vessels to carry the extra water taken

in by the growing root system and more bast-vessels to carry the extra food synthesized by the increasing number of leaves. The increased amount of wood in the stem also gives it strength to support the increasing weight of new branches and leaves. In

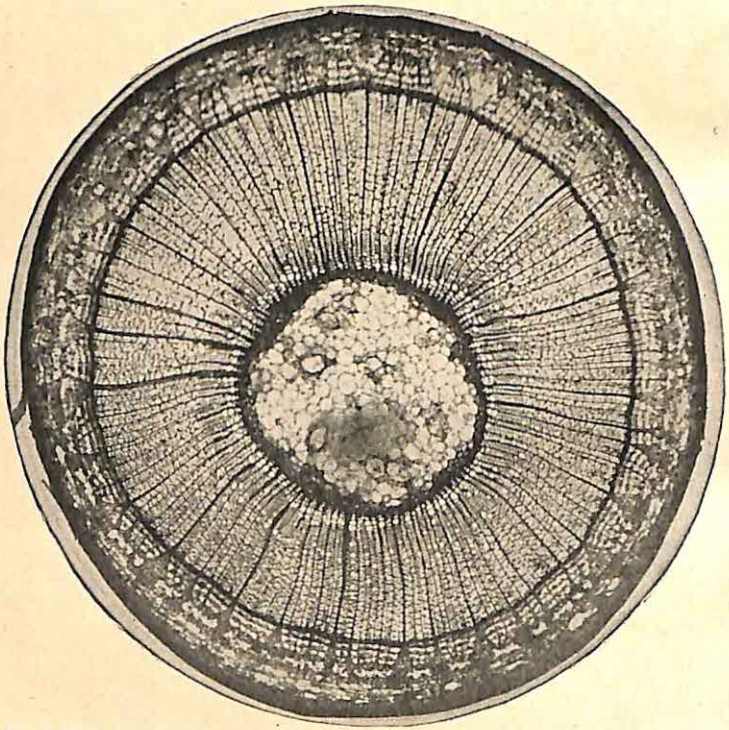


FIG. 96. Transverse section of young woody stem (dicotyledon), showing pith surrounded by complete cylinder of wood, with ring of cambium separating wood from bast. Notice the outer layer of bark and also the rays running through the wood (highly magnified).

short-lived, non-woody plants (which never produce very heavy branches) there is not the same need for a complete cylinder of wood—a ring of vascular bundles is sufficient. The new cells of this *secondary growth* are easily recognized in the cross-section of an older twig or trunk, because they all lie in rows, radiating from

the centre (see Fig. 96). Owing to this addition of layer after layer of wood it is often possible to estimate the age of a tree by counting the *growth-rings* on the cut surface of a tree-trunk (see Fig. 97). In wet seasons the wood grows rapidly and the vessels are thin-walled.

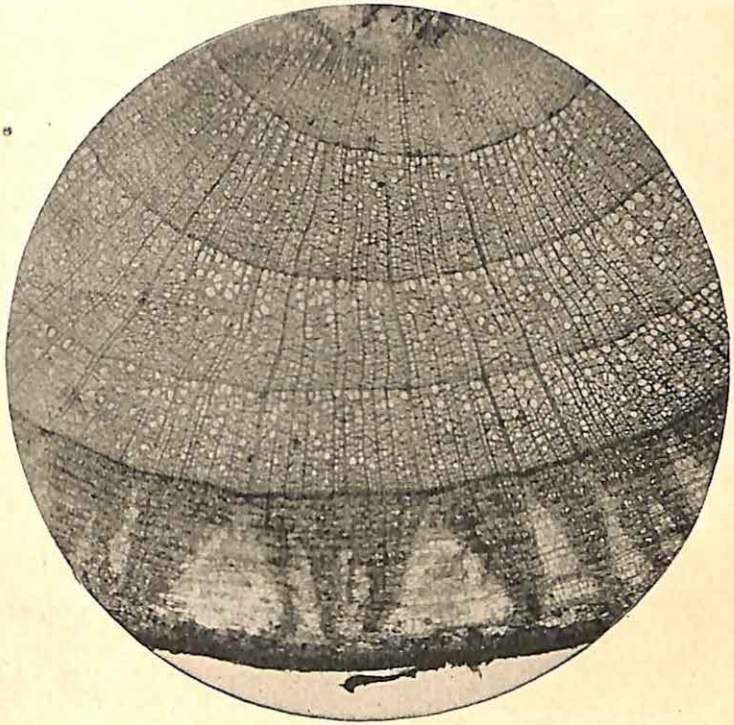


FIG. 97. Transverse section of older woody stem (dicotyledon), showing growth-rings and rays (highly magnified).

In dry seasons the wood grows very slowly and the cell-walls are thicker. In this way *growth-rings* are often seen on the cross-section of a tree-trunk, each ring representing the growth that took place in one wet season and one dry season. In temperate climates each ring usually represents one year's growth, but in tropical climates with more than one well-marked wet season each

year, there may be more than one ring formed each year, or growth may be so uniform that no distinct rings are visible.

The tubular structure of its vessels enables wood to resist a maximum bending force with a minimum of weight. The same is

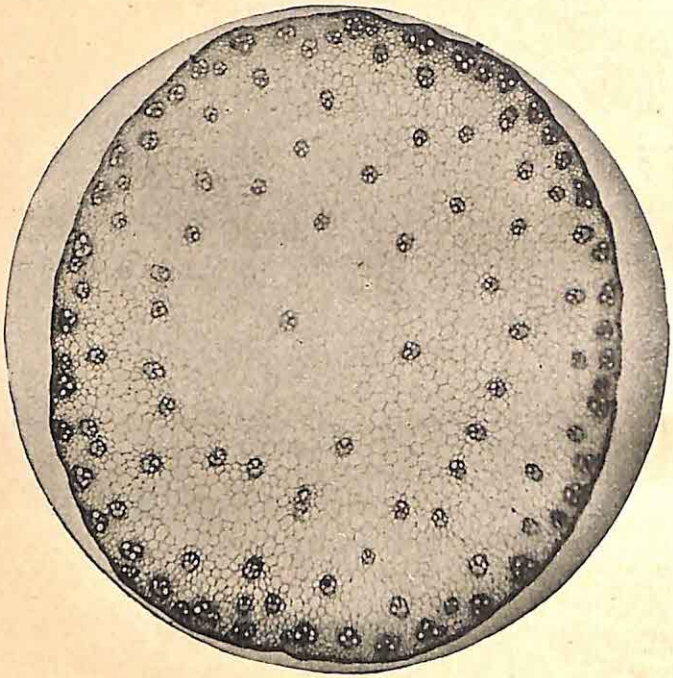


FIG. 98. Transverse section of maize stem (a monocotyledon), showing scattered vascular bundles (highly magnified).

true when we cut down a tree and use the timber* for making beams in a building. In the stems of monocotyledons no complete cylinders of wood and bast are formed, the vascular bundles remaining separate, so that the stems of monocotyledon trees are usually fibrous, and the timber (which is a mixture of wood and bast) is of little use for woodwork. The vascular bundles of monocotyledon stems have *no cambium*; hence there is no growth in

thickness; and the stems of such plants, e.g. Palms, are of practically the same diameter throughout their length (see Figs. 98 and 99).

Although the structure of an old thickened root is like that of an old thickened stem, a young root has a different internal structure

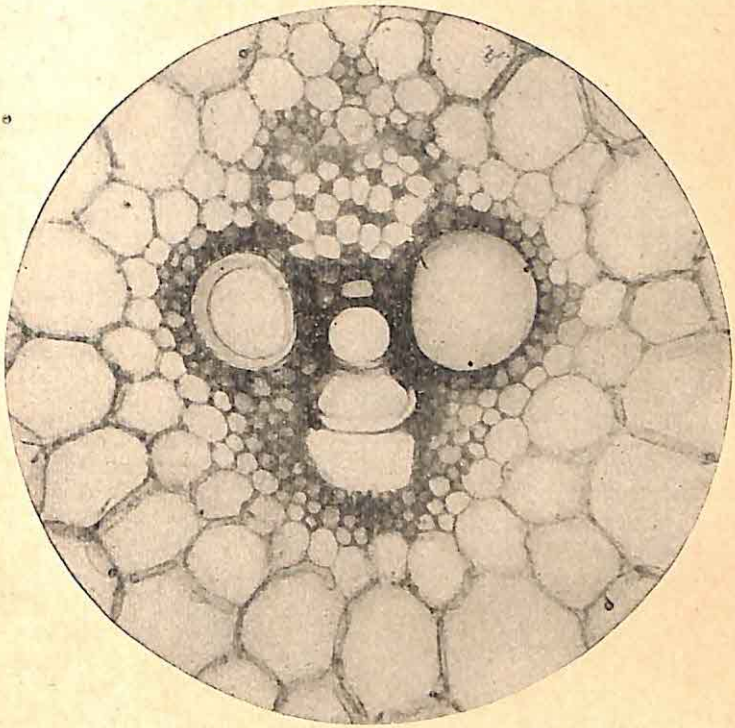


FIG. 99. More highly magnified transverse section of one of the vascular bundles shown in Fig. 98. The bundle contains xylem and phloem, but no cambium. (Compare with Fig. 95.)

from that of a young stem (see Fig. 100). In young roots there are separate bundles of wood-fibres and bast-fibres, arranged *alternately* to form a circle. In the root these fibrous bundles are nearer the centre than they are in the dicotyledon stem; thus they form a tough *central cylinder* (see Figs. 100 and 101). This is the best arrangement of strengthening tissue for resisting the *pulling strain**

that a root has to meet. (Stems usually have to stand a *bending* strain, and for this the best mechanical arrangement is to have the strengthening tissue nearer the *outside* of the stem.)

Where the root joins the stem, the wood-vessels of the root join

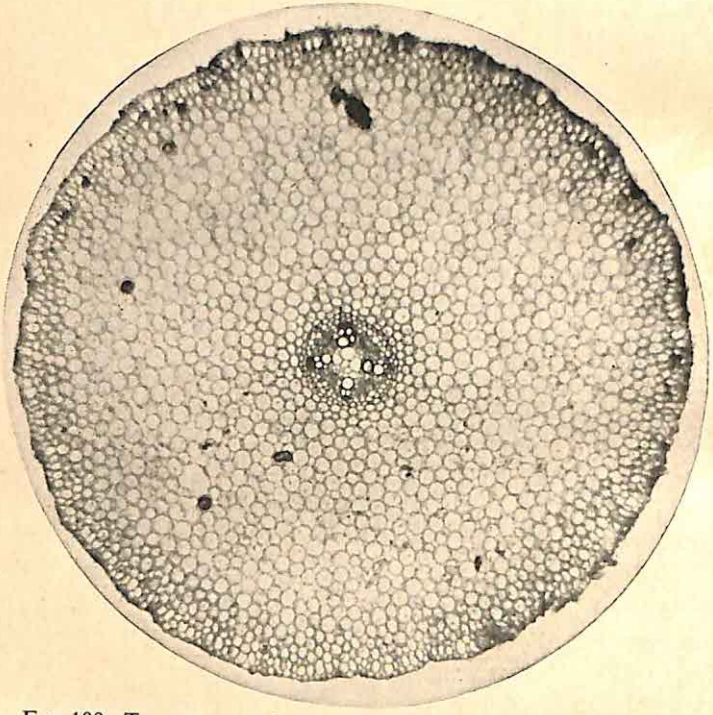


FIG. 100. Transverse section of young root of dicotyledon, showing central cylinder (highly magnified).

up with those of the stem, while the bast-vessels of root and stem are also continuous with each other. (Examine leaf, stem, and root sections under the microscope or micro-projector.)

THE PRACTICAL IMPORTANCE OF GROWTH MEASUREMENTS

Growth measurements are of great importance to biologists in many ways. In plants, many movements of stems and roots can be explained if we know something of the factors that cause

changes in rate of growth in length. In both animals and plants, growth *in weight* is affected by a great many factors—by temperature, light, food supply, by vitamins and hormones.

Agricultural research workers and medical scientists therefore

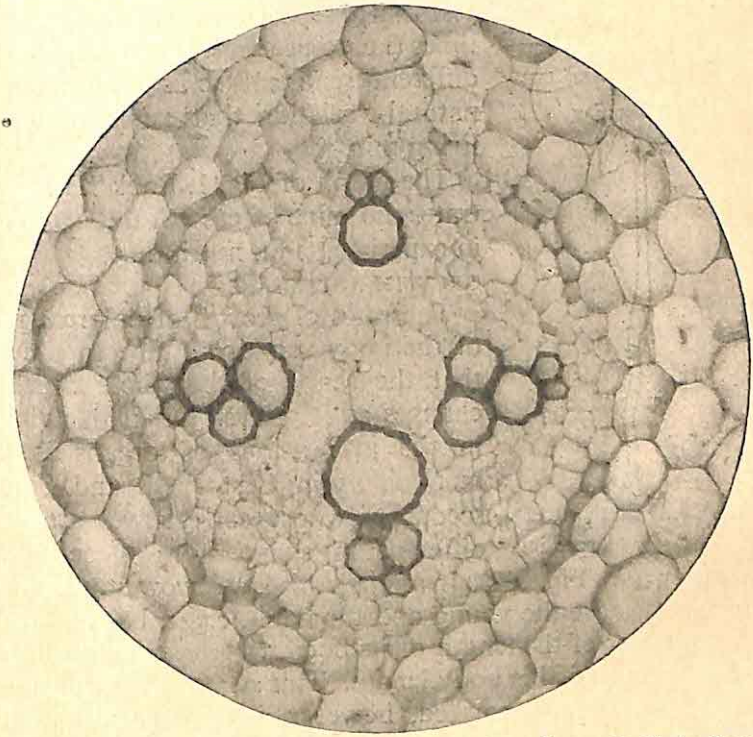


FIG. 101. More highly magnified transverse section of central cylinder from Fig. 100, showing bundles of wood-vessels (thick-walled) alternating with bundles of bast-vessels (thin-walled).

measure growth in order to discover whether a plant or animal has been given the most favourable conditions. As we have seen, growth for this purpose is best measured in plants by the change in *dry-weight*. But with animals, which do not undergo such marked changes in water-content as do plants, it is usually sufficient to measure growth by measuring changes in *fresh-* or *live-weight*.

This method is simple and it has one great advantage: each series of growth measurements can be carried out on the same animals instead of on a number of different samples (as with plants). Nevertheless, in all work of this kind it is essential to repeat each experiment several times (i.e. to use several animals for each different treatment) so as to allow for the natural variation that exists among all living things.

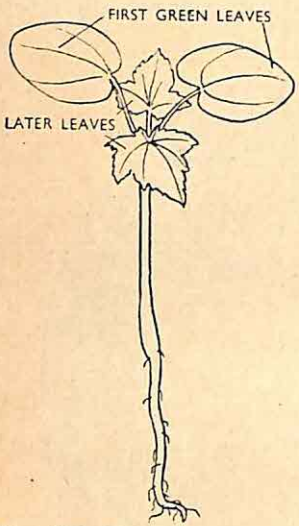


FIG. 102. Different shapes of leaves in castor oil seedling.

The graph of Fig. 1, Book III A, illustrates the value of growth measurements in experiments on nutrition.

GROWTH AND SHAPE

The rates of growth of different parts of the body vary from time to time, so that the shape of the whole organism changes as it grows. Often this change in the proportions of organs is well marked, as in Castor Oil and Squash seedlings, whose early leaves are quite different in shape from those of the mature plant (see Fig. 102). In animals, also, we know how different in proportions are the young and the adult, e.g. a baby's head is about a quarter of the length of its body, whereas a man's head is only one-seventh of his body-length. This change in form during development is most marked, however, in insects and amphibians. Here the animal passes through several distinct stages in its development and we speak of its *metamorphosis* (see Book Three, Chap. III).

REPRODUCTION IN PLANTS AND ANIMALS

We have seen that one of the chief characteristics that distinguishes living things from non-living matter is that *living things can reproduce themselves* while non-living things cannot. Except for some of the simplest one-celled organisms like Amoeba and Bacteria,

all living things have a limited length of life and are finally overtaken by *death*. Many plants and animals die by accident before they grow old, e.g. they are often eaten by animals, they often die from lack of food, and they are often killed by parasites. But even those living things that escape death by accident finally die of old age.

With the exception of the simplest one-celled organisms, every *individual* living thing must die, but there is a continuous production of new individuals to replace the dead ones and thus maintain the *race*. These new individuals develop from small pieces of living matter that have *separated from other individuals* in the process of *reproduction*.

The larger plants and animals are made up of numerous cells, and each cell was produced by the division of another previously existing cell. The new cells remain joined together and become built up into tissues and organs.

In very simple one-celled organisms like Bacteria and Amoeba the new cells separate from each other and form new individuals. In such cases, simple *division* is a very efficient means of *reproduction*. For example, when there is plenty of suitable food a bacterial cell divides into two new cells every 30 minutes; in 10 hours it can produce about a million new bacteria. Such a very rapid increase is not often found in Nature because the growing cells soon become short of food, so growth slows down. But the possibility of such quick growth and multiplication explains the rapid decay of dead plants and animals and the rapid development of diseases caused by bacteria.

REPRODUCTION IN PLANTS

The simplest type of reproduction is by *division*: one individual splitting into two similar halves, which then grow up into two new individuals. This primitive type of reproduction is only possible in the simplest plants and animals and is commonest in one-celled organisms. For example, free-swimming one-celled plants usually reproduce in this way.

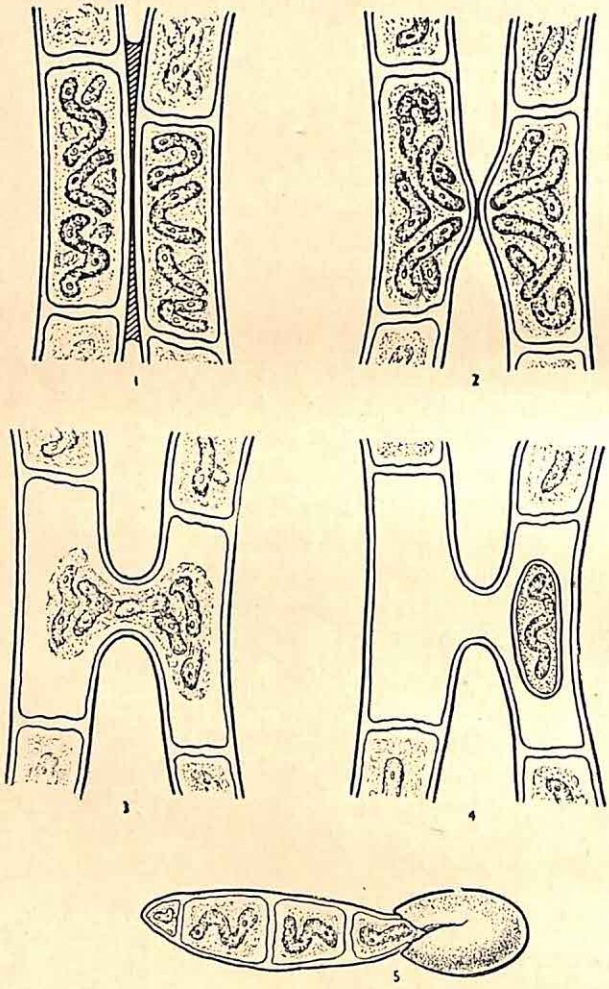


FIG. 103. Primitive sexual reproduction in Spirogyra (magnified). 1. Threads sticking together. 2. Cells beginning to unite. 3. Cell-contents fusing together. 4. Resting-stage. 5. New plant germinating (after Hatfield).

In fully-grown cells of Spirogyra the nucleus divides into two equal halves and a cell-wall grows across the cell, forming two 'daughter-cells',[1] each of which is a separate living unit although they remain attached end to end. This primitive type of reproduc-

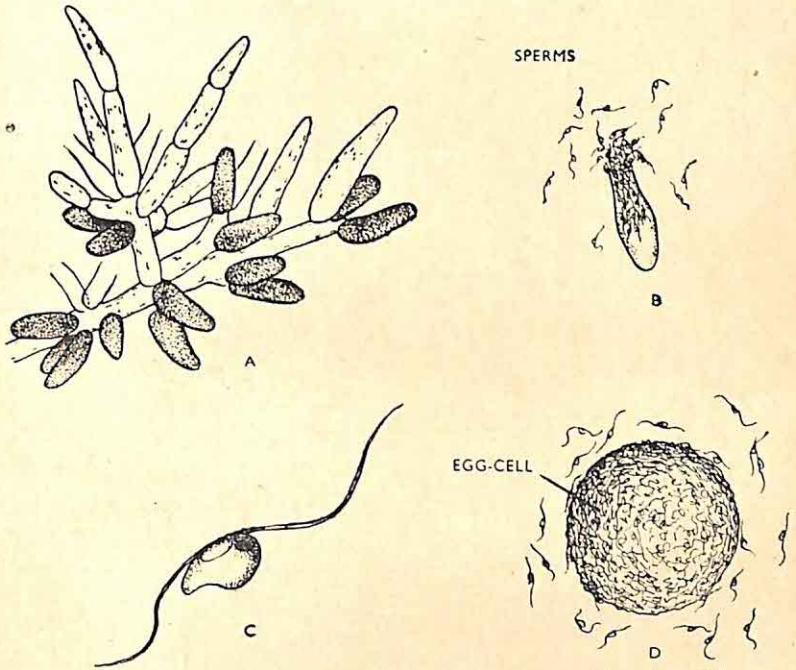


FIG. 104. Fertilization in a sea-weed (magnified). A. Branching thread from male reproductive organ. B. Sperms escaping. C. A single, free-swimming sperm (still more highly magnified). D. An egg-cell surrounded by sperms (after Hatfield).

tion, where the new individuals develop from one single parent and where there is no pairing or *fertilization*, is called *vegetative reproduction* or *asexual reproduction*.

Spirogyra also shows a different type of reproduction in which *two cells* take part. When two threads of Spirogyra lie side by side two opposite cells may send out projections of the cell-wall that

[1] In describing cell-division, it is convenient to use the terms 'mother-cell' and 'daughter-cell', but you must remember that these terms do not refer to 'female' cells.

meet and finally form a passage uniting the two cells (see Fig. 103.) The contents of one cell pass through this passage and fuse with

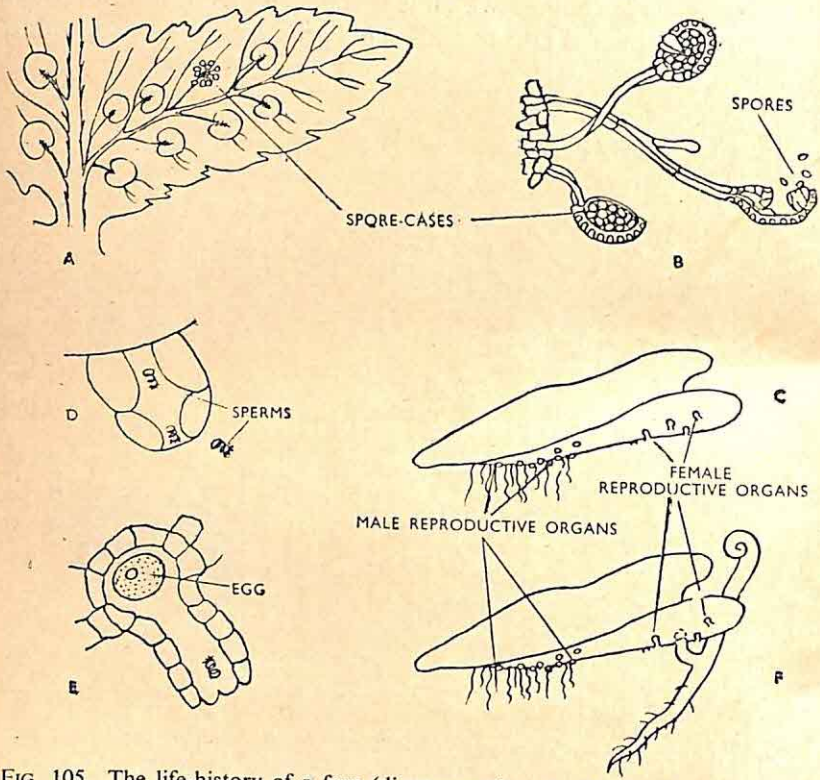


FIG. 105. The life-history of a fern (diagrammatic). A. Underside of fern leaflet showing groups of spore-cases (one without its cover). B. Three spore-cases, magnified, with one opening and setting free its spores. C. A *spore plant*, with male and female reproductive organs on its underside (also thread-like 'roots'). D. A male reproductive organ with sperms escaping (highly magnified). E. A female reproductive organ containing an egg-cell and with one sperm entering (highly magnified). F. Fertilized egg-cell developing into a new *fern-plant* (after Holmes and Gibbs).

the contents of the other cell. This single reproductive cell, formed by the fusion of two Spirogyra cells, grows a thick protective layer of cellulose and rests in this condition until external conditions become favourable for growth, when it grows into a new thread of

Spirogyra. This type of reproduction, resulting from *fertilization* (in which *two cells unite to become one cell*), is called *sexual reproduction* (although in the case of Spirogyra we cannot say which of the two cells taking part is the male cell and which the female cell—so that Spirogyra shows only primitive sexual reproduction).

Seaweeds show a more advanced type of sexual reproduction. There are two quite different kinds of reproductive cells: (*a*) very tiny, active, free-swimming *sperm-cells*, and (*b*) passive* *egg-cells*, several thousand times bigger than the sperm-cells (see Fig. 104). The male sperm-cells swim through the surrounding water to the female egg-cells and when one sperm-cell enters an egg-cell and their nuclei fuse together, the *fertilized egg-cell* can develop into a new seaweed plant.

Reproduction in Ferns and Mosses is rather complicated as there are two different methods of reproduction that take place alternately. On the under-sides of many fern leaves can be seen small brown or black patches (see Fig. 105). These cover groups of *spore-cases* that contain numerous microscopic reproductive cells called *spores*. The ordinary, conspicuous fern-plant, therefore, is the *asexual generation*, producing spores without any pairing or fertilization. These spores, however, *do not grow directly into new fern-plants*, but they produce a small, flat, green *spore-plant* (often found growing on damp earth under the parent fern-plant). This tiny spore-plant is the *sexual generation* since it produces egg-cells and sperm-cells. The sperm-cells swim through a drop of water tc the ovaries, and when a sperm-cell fuses with an egg-cell, the fertilized egg-cell develops into a new *fern-plant*. Hence there is a regular *alternation of generations*—the asexual fern-plant producing spores and the sexual spore-plant producing female egg-cells and male sperm-cells.

Reproduction in *flowering plants* is adapted to life on dry land. The sperm-cell is carried through the air enclosed in a pollen grain and the last part of its journey to the egg-cell takes place inside the pollen tube. Pollen is carried from stamen to stigma

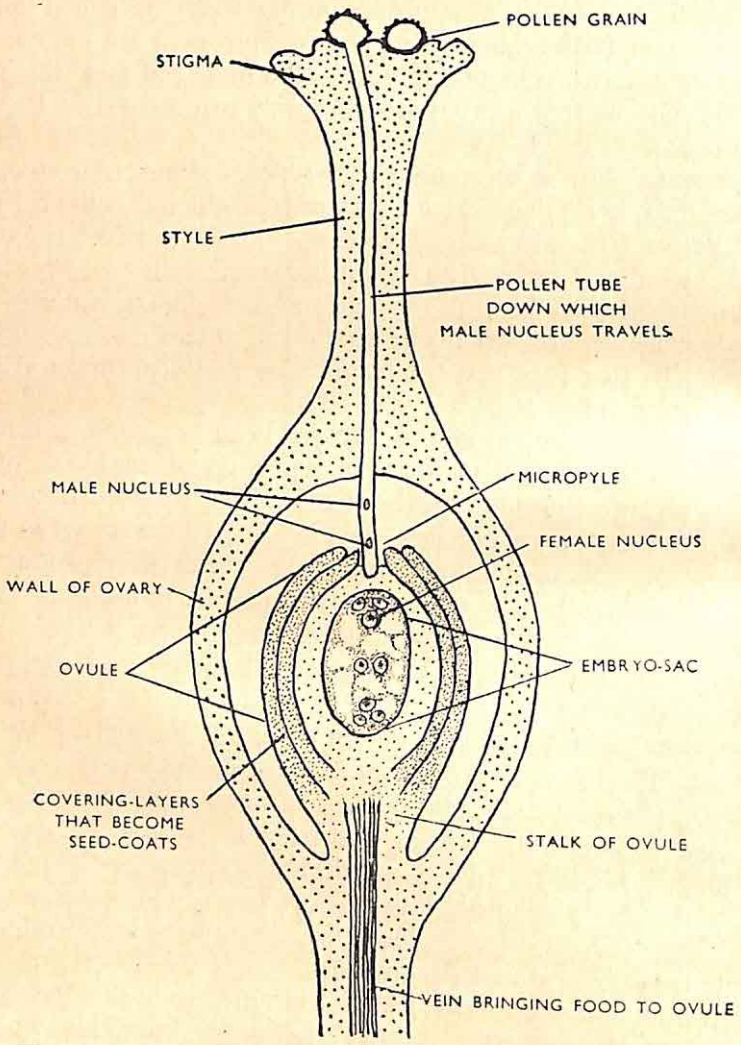


FIG. 106. Fertilization in a flowering plant (diagrammatic and much enlarged—after Hatfield).

either by *wind* or by an *insect* visiting the flower in search of food. Both sperm-cells and egg-cells are protected throughout the process leading up to fertilization, and afterwards the young embryo is protected and nourished inside the seed and receives a food store from the parent plant. The seed, when ripe, becomes dry and is thus able to withstand unfavourable conditions, such as cold or drought. In most flowering plants the parent plant scatters the seeds and gives them a better chance of finding a suitable spot in which to grow. Thus the seed is an organ of *reproduction*, of *survival* from year to year, and also of *dispersal** (see Fig. 106).

VEGETATIVE REPRODUCTION IN FLOWERING PLANTS

The simplest one-celled plants can reproduce *by division*. The more complicated and specialized higher plants cannot split into two similar halves in this way, but in many flowering plants a suitable piece cut from one plant can grow into a new and complete plant. This method of reproduction from some portion of the plant other than the seed is called *vegetative reproduction* because only *vegetative* parts of the plant are used (as distinct from *seeds*, which are derived from the male and female—or *sexual*—organs of the plant).

Vegetative reproduction is widely used by gardeners and agriculturists, particularly in tropical countries. *Underground stems* are often used as the starting-point of a new plant, e.g. Onions, Lilies, and Crinums can be grown from *bulbs*, Yams, English Potatoes, and Sweet Potatoes can be grown from *tubers*, Caladiums can be grown from *corms*, Ginger and Arrowroot can be grown from *rhizomes*, Bananas can be grown from *root-suckers*, Pineapples can be grown from *stem-suckers* (as well as from root-suckers). *Stem-cuttings* will often root readily and develop into new plants.

Another method of vegetative reproduction is by *grafting*.† In grafting, a shoot of the required plant, e.g. a good variety of Orange or Mango, is placed with its cut surface in contact with the cut surface of a rooted plant (called the 'stock'†) so that the two

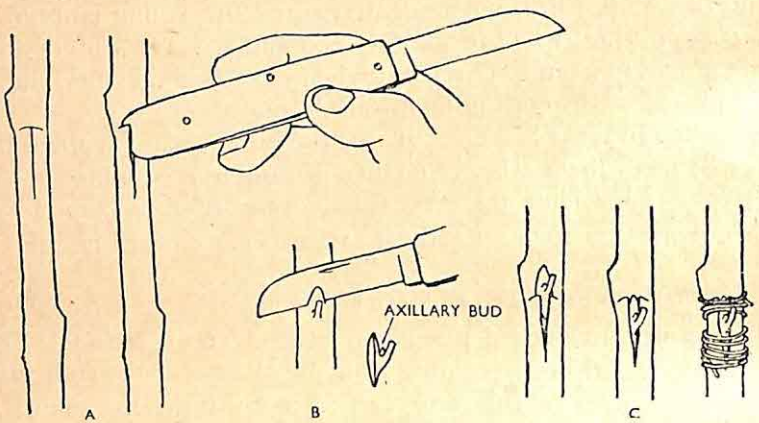


FIG. 107. Bud-grafting. A. Preparing slit in bark of stock. B. Removing bud with patch of bark, bast, and cambium from plant to be grafted on to the stock. C. Fixing bud-graft in slit on stock (after Irvine).

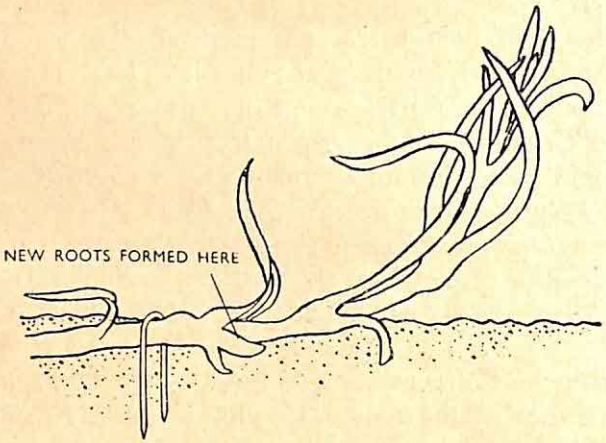


FIG. 108. Vegetative reproduction by layering (after Gurr).

cambium layers grow together. The 'stock' supplies a strong and hardy root-system while the grafted shoot will produce fruits of the same superior quality as the parent tree from which the graft was cut. In tropical countries, however, the commonest method

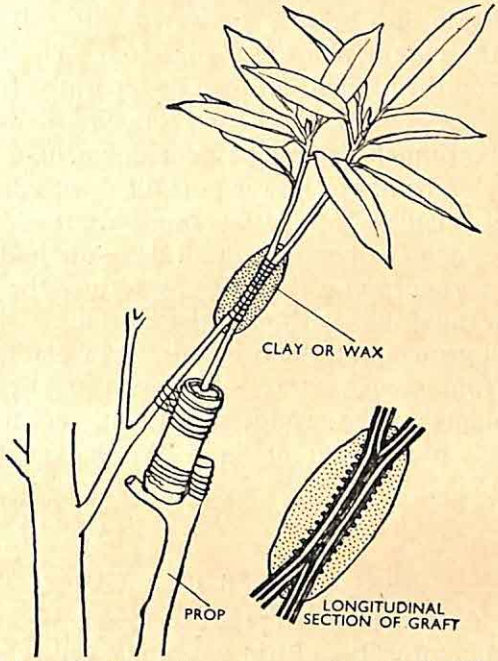


FIG. 109. Stem of sweet mango being grafted 'by approach' to semi-wild mango stock in bamboo 'flower-pot' (after Irvine).

of grafting is *budding* (or *bud-grafting*), a single bud of the desired plant being grafted on to a suitable rootstock (see Fig. 107). It is, however, only possible to graft together plants of the same or closely related kinds. This method is used for producing new rubber trees from high-yielding parents. In this way, also, a common red Hibiscus can be made to bear several different colours and varieties of Hibiscus flowers by bud-grafting. Similarly, a

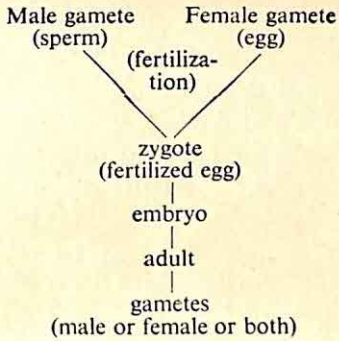
Shaddock† (Pumelo) tree can be bud-grafted with any other citrus fruit, e.g. Orange, Lime, Lemon, or Grapefruit, so that it is possible to have one tree bearing all these different fruits. An alternative method of grafting 'by approach' is shown in Fig. 109.

Another method of vegetative reproduction is *layering*.† In this method a portion of a branch is covered with soil while it is still attached to the parent plant (see Fig. 108). The buried portion grows roots and the branch can then be separated from the parent plant. (This is a good method for producing new Bougainvillea plants.) If the branch cannot be bent down to the ground it is passed through a bamboo 'flower-pot' filled with damp soil. This latter method is called 'gootee layering'† or '*marcottage*'.†

The great practical advantage of all these methods of *vegetative* (or *asexual*) *reproduction* is that they reproduce the characteristics of one parent only, just as though the cutting or other vegetative part were still growing on the original parent plant. When plants are grown from seeds there is always a possibility that new varieties of plants will be produced. In fact, very few of the seedlings are exactly like the parent plant. We shall explain the reason for this later.

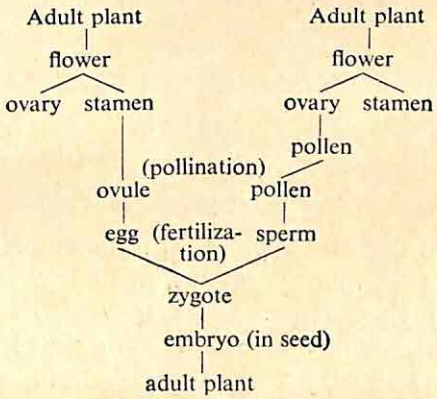
SEXUAL REPRODUCTION

Most animals and plants reproduce by a process of *fusion* between two different cells to form one single cell, which then grows into a new organism. The fusing cells are called sex-cells or *gametes*.† They are usually different in size and behaviour and we call one the *male gamete* or *sperm*, and the other the *female gamete* or *egg* (*ovum*). The adult plant or animal may produce (in its *sex-organs*) either male cells only, or female cells only, or both kinds. Usually, however, in *sexual reproduction*, an organism grows from a cell that comes from *two* different parents. The fusion of the gametes is called *fertilization* and the fused cells form one *fertilized egg cell* (or *zygote*†), which grows into a new organism. This process may be summarized in the following way:

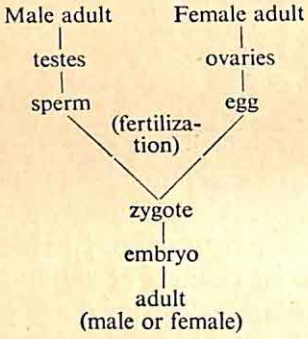


Sexual reproduction, which takes place in practically all organisms (an important exception being Bacteria), is of great importance in *evolution*, because it means that an individual can inherit some of the characteristics of its two parents. We shall learn later how this increases the chances of variation and change.

Although some plants are uni-sexual (having stamens and pistils in separate flowers) most of them produce both male and female gametes. Nevertheless it remains true that two different parents are usually concerned in sexual reproduction. The sex organs are produced in the flowers, and the following diagram summarizes the reproductive cycle:



Most of the higher animals are uni-sexual (i.e. the individuals are either male or female) and the two sexes exist in roughly equal numbers. The male animal produces *sperms* in special male sex-organs (*testes*), and the female animal produces *eggs* in its *ovaries*. The two gametes are brought together during the mating or pairing of the sexes and the fertilized eggs develop into either male or female adults.



REPRODUCTION IN ANIMALS

The simplest one-celled animals, like the one-celled plants, divide into two equal parts when they reach their full size, and each half then develops as a new individual. Amoeba and the Bacteria reproduce in this way by simple division into two (see Fig. 110), and

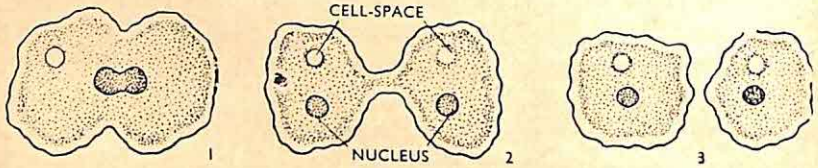


FIG. 110. Amoeba reproducing by division.

we have explained that this type of reproduction is called *asexual reproduction*, because there is no fusion of two cells or any exchange of cell-contents between two cells.

Paramecium, like Amoeba, usually reproduces asexually, one individual dividing into two. Under favourable conditions this

type of reproduction may take place at intervals of less than one day. Occasionally, however, a primitive type of sexual reproduction takes place in Paramecium. Two individuals pair together, 'mouth' to 'mouth', so that the protoplasm of both cells becomes

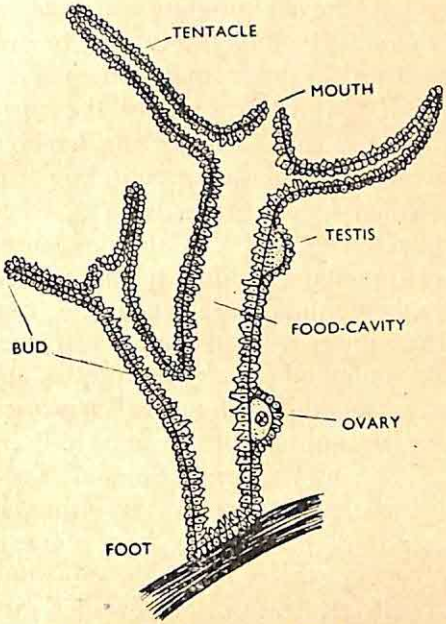


FIG. 111. Longitudinal section of Hydra (magnified).

continuous, and there is an exchange of parts of their nuclei. The two individuals then separate and afterwards reproduce asexually, by division.

Hydra shows both types of reproduction—asexual and sexual. Although Hydra does not reproduce naturally by simple division, if it is cut in half each half can grow the missing parts, thus producing two new animals. The natural method of *asexual reproduction* in Hydra is by *budding*. A small swelling, or *bud*, arises on the side of the animal and grows out into a new Hydra. When the bud has developed a 'mouth' and can get its own food, it separates from its parent and lives independently (see Fig. 111).

Hydra can also reproduce *sexually*, by the fusion of special reproductive cells—egg-cells and sperm-cells. Tiny swellings are formed on the outside of the body, near the mouth end. These become the *male reproductive organs*, or *testes*, producing *sperms*, which are set free into the surrounding water when the ripe testis bursts open. The *female reproductive organs*, or *ovaries*, are larger swellings near the 'foot' of the animal, and each ovary produces a single, large *egg-cell* that is exposed to the surrounding water when the ovary becomes ripe. The sperms, which are shaped like tiny tadpoles, swim through the surrounding water by moving their 'tails', and when a sperm reaches a ripe egg-cell it enters, producing a fertilized egg that can develop into a new Hydra (usually after a resting-stage inside a horny, protective covering).

Most of the higher animals are *uni-sexual*, but in *Hydra* and some of the other lower animals, e.g. earthworms and snails, one animal is both male and female, both testes and ovaries being present in the same animal. Such animals are *bisexual* (or *hermaphrodite*). In *uni-sexual* animals there can be only *cross-fertilization* since the pairing egg and sperm come always from different individuals. In *bisexual* animals, however, either *cross-fertilization* or *self-fertilization* may take place, according to whether the fertilizing sperm comes to the egg from the same or from a different individual. Such self-fertilization, however, is rare, and we shall see that the earthworm and the snail, for example, although bisexual, are always cross-fertilized. (It is interesting to note that while the highest animals are always uni-sexual, the highest plants are nearly all bisexual, having stamens and pistil both present in the same flower.)

In *earthworms* both sexes are combined in one bisexual animal, yet only cross-fertilization can take place. This makes the reproductive system very complicated. In the Indian earthworm (*Pheretima*) there are two pairs of *testes*, tiny white bodies enclosed in two pairs of *testis-sacs* lying ventrally below the food-canal in the 10th and 11th segments (see Fig. 112). The *sperm-cells* produced by the testes are passed into two pairs of *sperm-sacs* lying

in segments 11 and 12 where they complete their development. Two pairs of very fine *sperm-ducts* convey the sperms to a pair of male reproductive openings in segment 18. The female reproductive

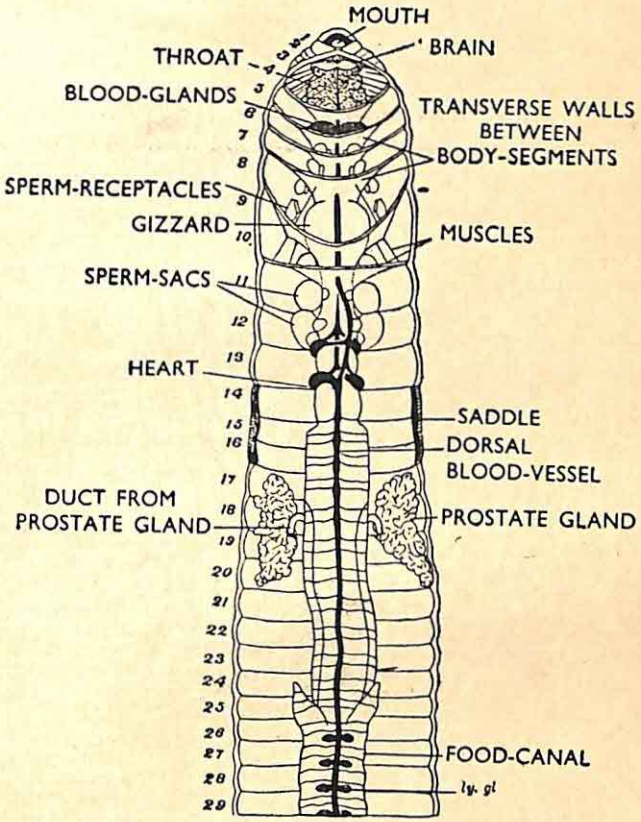


FIG. 112. Indian earthworm dissected from dorsal surface to show contents of anterior portion of body-cavity (after Bahl).

organs are a pair of tiny *ovaries* lying ventrally below the food-canal in segments 13 and 14. The *egg-cells*, when ripe, are conveyed through a pair of very fine *oviducts* to the female reproductive opening in segment 14.

Besides these male and female organs, the earthworm has

several pairs of *sperm-receptacles* in segments 6–8, in which are stored 'foreign' *sperms received from another worm.* In order to

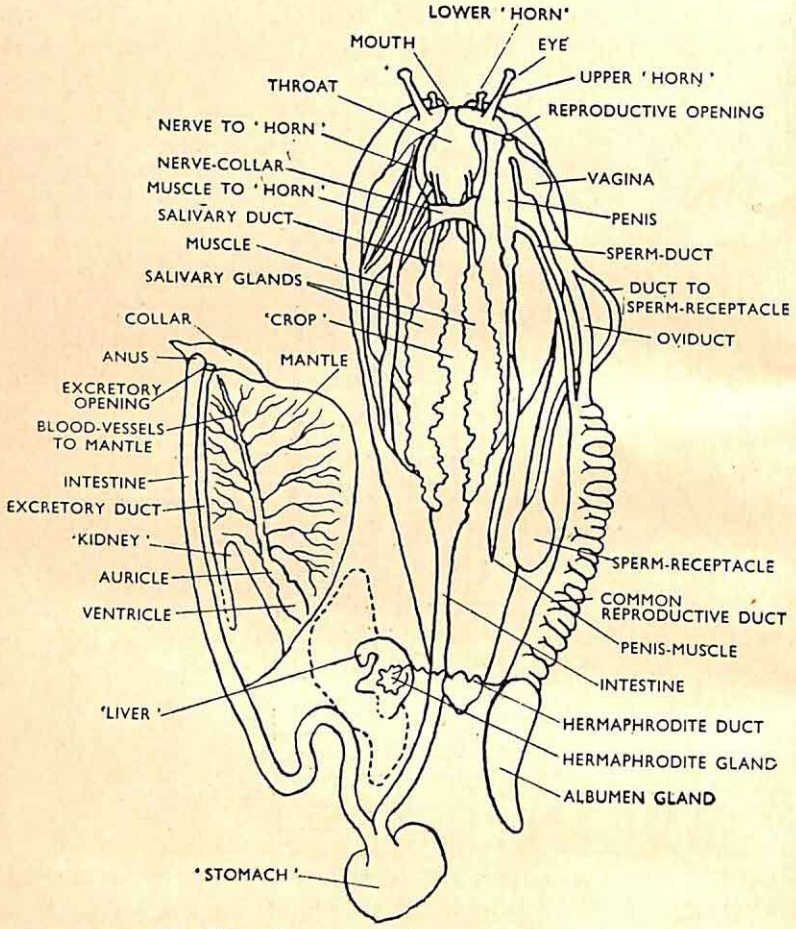


FIG. 113. Dissection of giant snail (dorsal view).

exchange sperms, two coupling worms place themselves together in a head-to-tail position so that the male reproductive openings of one worm are opposite the external openings of the *sperm receptacles* of the other worm. Sperms are exchanged and the

two worms then separate. Fertilization takes place later. A band of elastic skin is formed from the 'saddle' and the worm withdraws backwards out of this band. As the band of skin passes over the female reproductive opening, egg-cells are passed out into it, and as it passes over the external openings to the sperm-receptacles it receives sperms (which came originally from another worm). Fertilization then takes place, and as soon as the band of skin is clear of the anterior end of the worm it closes up to form a horny *egg-case*. In each egg-case only one fertilized egg usually develops into a young worm. It is remarkable to find such complicated arrangements for reproduction in an animal whose other body-systems are comparatively simple, but all these complications make self-fertilization impossible and result in cross-fertilization.

Snails, like earthworms, are bisexual but not self-fertilizing. Both egg-cells and sperm-cells are produced by the same *hermaphrodite gland* (or *ovo-testis*†), which is embedded in the liver near the apex of the shell (see Fig. 113). The eggs and sperms then pass down the wavy, white *hermaphrodite duct*, leading to the large *albumen-gland*, where the eggs receive a layer of albumen (protein) which later serves as food for the developing embryo. After leaving the albumen-gland the eggs and sperms travel along separate ducts—an oviduct and a sperm-duct—and finally the male and female ducts join again to open to the exterior at a *common reproductive opening* on the right side of the snail's head.

Before the oviduct and the sperm-duct finally join together, the last part of the oviduct (the *vagina*) becomes wide and thick-walled, and from its posterior end arises a long tube ending in a sac—the *sperm-receptacle*—in which are stored 'foreign' sperms received from another snail. The lower end of the sperm-duct leads into a muscular tube—the *penis*—that can be pushed out through the reproductive opening and into the reproductive opening of another snail. Mating snails *exchange sperms* by placing their reproductive openings in contact, and sperms from one animal are transferred to the sperm-receptacle of the other, where they are stored. The two animals then separate. As eggs pass

down the lower end of the oviduct they are cross-fertilized by 'foreign' sperms from the sperm-receptacle and the fertilized eggs are enclosed in shells and laid in a hole in the soil. As in the earthworm, the snail's reproductive arrangements are very complicated owing to the difficulty of ensuring cross-fertilization in a bisexual animal.

Insects are uni-sexual and sperms are passed from the male into the body of the female, i.e. there is *internal fertilization*. This type of fertilization is characteristic of all land animals, e.g. snails, insects, reptiles, birds, and mammals, since external fertilization is only possible in animals that breed in water, e.g. most fishes and amphibians. The reproductive system of insects is rather complicated and difficult to make out except in the larger insects. In the male cockroach, for example, the testes are a pair of tiny oval bodies lying in the 5th and 6th segments of the hind-body. They can only be seen in young males since they waste away before the animal is fully grown. Sperms are passed from the testes down a pair of sperm-ducts to the sperm-sacs, which consist of a number of small, blind-ended tubes, bunched together in the shape of a mushroom. (This mushroom-shaped gland can be seen in adult male cockroaches and was formerly mistaken for the testis.) A single duct leads from the sperm-sacs to the male reproductive opening, which is just below the anus. In the female cockroach there is a pair of ovaries in the posterior part of the hind-body (see Fig. 44). Each ovary, right and left, consists of a bunch of eight egg-tubes, and, owing to the eggs inside, each egg-tube looks like a string of beads. The eggs pass from the egg-tubes into the oviducts, which join together before opening at the female reproductive opening into the boat-shaped structure at the posterior end of the hind-body. Just above this female reproductive opening is the opening to the small, tubular sperm-receptacle, in which are stored sperms received from a male cockroach during pairing. This store of sperms is used to fertilize the eggs as they are laid. As the eggs pass out of the female reproductive opening they are fertilized by sperms from the sperm-receptacle, and a

number of fertilized eggs (14–24) is then packed in a horny egg-case. The female cockroach carries this egg-case, projecting from the boat-shaped structure, until she finds a suitable place to leave it.

Fishes are uni-sexual, i.e. each individual is either a female with ovaries, producing eggs, or a male with testes, producing sperms. In the bony fishes, e.g. Carp, there is *external fertilization.* The female fish discharges a very large number of eggs and the male fish swims over the eggs and discharges sperms into the surrounding water. This uncertain method of fertilization is very wasteful, and enormous numbers of eggs and sperms are produced to compensate for the great wastage. A female cod-fish, for example, may produce over four million eggs each breeding season, of which only a very small proportion will ever be fertilized and develop.

In gristly fishes, e.g. sharks, this great wastage is avoided by *internal fertilization.* The sperms are introduced by the male directly into the female oviduct where they meet the eggs. In some gristly fishes the fertilized egg is then enclosed in a horny egg-case, together with a large supply of food, before being passed out into the surrounding water to 'look after itself'. In some other gristly fishes, e.g. the Indian shark (*Scoliodon*), after internal fertilization, the fertilized eggs attach themselves to the lining of the lower part of the oviduct and the developing embryos obtain their food and oxygen from the parent fish. When the young fish are well developed and several inches in length, they are born alive. This type of reproduction is clearly much less wasteful than the method of passing out enormous numbers of eggs and sperms into the surrounding water and leaving them without any further attention —without any *parental care.*

In *frogs* fertilization is *external.* The male frog has a pair of testes: pale yellow, oval bodies attached to the anterior end of the kidneys. A number of fine ducts convey sperms from each testis through the corresponding kidney, whence they pass down the kidney-duct (which carries both sperms and urine), and are then stored in a sperm-sac ready to be poured out through the

cloaca, over the eggs laid by the female frog. The female frog has a pair of ovaries, lying in the same position as the male testes, and consisting of irregular masses of black and white egg-cells. The

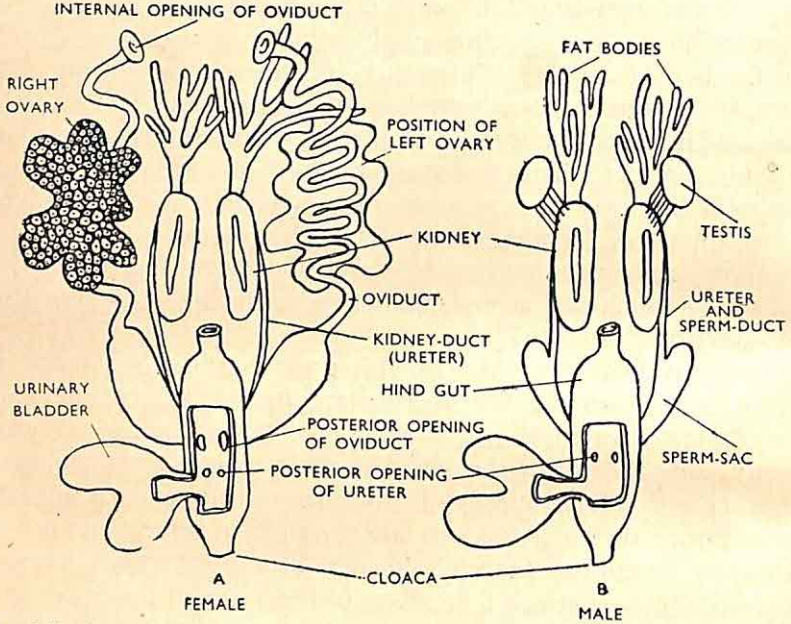


FIG. 114. Reproductive (and excretory) organs of frog (after Hatfield). A. Female frog. B. Male frog. The wall of the cloaca has been cut away (and the bladder displaced to one side) to show hinder openings of excretory and reproductive ducts.

egg-cells pass down through the right and left oviducts, long, coiled, white tubes, into the cloaca (see Fig. 114).

Frogs mate in water, and as the eggs are discharged from the cloaca of the female, the male pours sperms over them, so that fertilization is *external*. As the eggs reach water their external slimy coat swells up and sets to form a thick layer of stiff jelly. Through this a swimming male sperm makes its way, to fuse with the female egg-cell. In Book Three, Chap. III, we discussed the development of the frog from such a fertilized egg.

Birds, like the other higher animals, are uni-sexual, and in many

cases there are striking external differences in the colour and arrangement of the feathers and also in the size of the body in the two sexes. The reproductive organs of birds follow the usual vertebrate plan except that there is only one ovary and oviduct (the

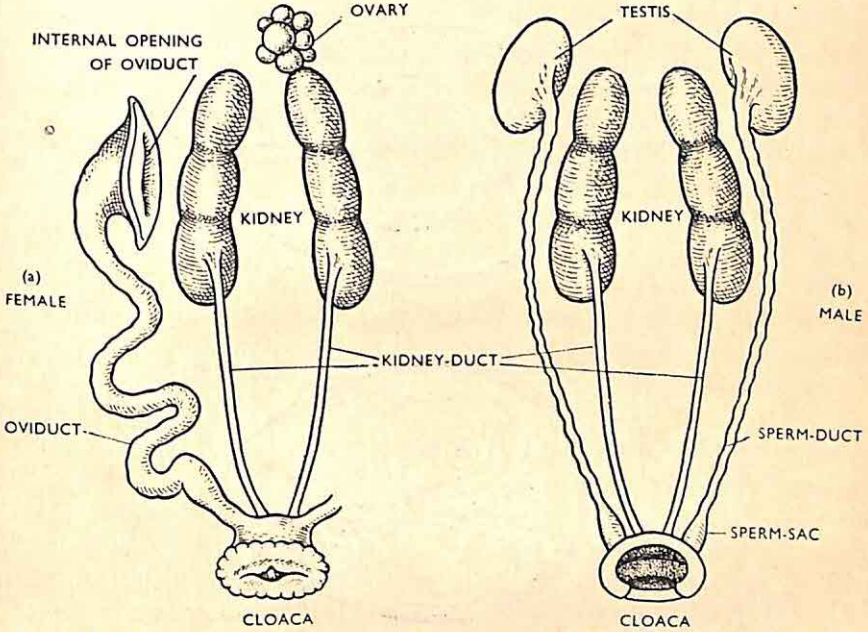


FIG. 115. Reproductive (and excretory) organs of (a) female and (b) male bird (after Thomson).

left) in the adult female (see Fig. 115). In an adult female domestic fowl, during the breeding season, the ovary consists of a bunch of small, rounded egg-cells, situated on the dorsal wall of the body-cavity, just anterior to the kidneys. A fully-developed egg-cell, about an inch in diameter, enters the oviduct, and, if the female bird has received a supply of sperms from a male bird during pairing, fertilization takes place *in the upper part of the oviduct*, a sperm fusing with the egg. As the egg passes down the oviduct, it first receives a thick coating of *albumen* (egg-white = about 13 per cent. protein and 86 per cent. water), then the transparent

shell-membrane, and finally the *shell* itself, consisting of a porous layer of calcium carbonate. About 24 hours after leaving the ovary the egg is laid, passing out through the *cloaca*, the common reproductive and excretory opening. In the adult male bird, during

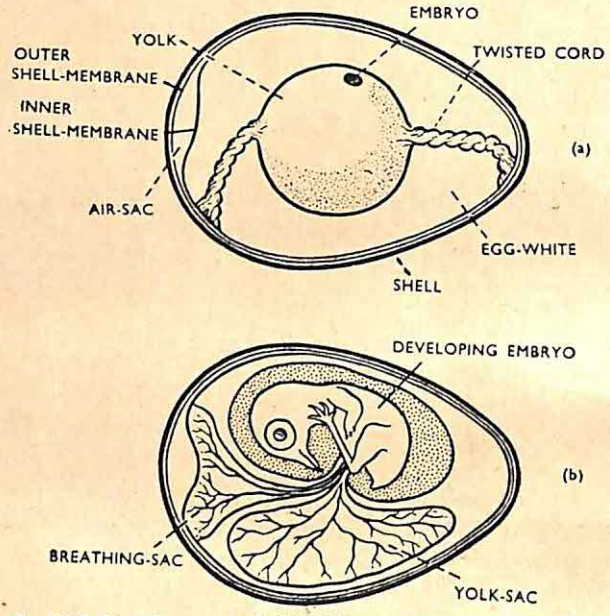


FIG. 116. Development of chick (diagrammatic). (a) Fertilized egg when newly laid. (b) After 14 days under sitting hen.

the breeding season, the testes are two oval, light-coloured bodies, situated on the dorsal wall of the body-cavity just anterior to the kidneys, and ripe sperms are passed down the right and left sperm-ducts to a pair of sperm-sacs near the cloaca (see Fig. 115).

During pairing, sperms are passed from the male to the female cloaca, along with some liquid, and the sperms travel up the oviduct and fertilize the egg-cells. Fertilization, therefore, is *internal*, as in other land-animals.

The development of the chick. The new-laid egg, as shown in section in Fig. 116, consists of (*a*) a porous *shell* through which

exchange of gases can take place between the developing embryo and the atmosphere, (*b*) a double *shell-membrane* forming an *air-chamber* at the broad end of the shell, (*c*) the *egg-white*, supplying food and water to the developing embryo, and (*d*) the *egg-cell*, consisting of the tiny *embryo*—a white disc, about 5 mm. in diameter, always on the upper surface—and the large *yolk*, enclosed in a very thin membrane, supplying food for the developing embryo. (The egg-cells of birds are among the largest of all cells.) Identify all these parts in a new-laid egg that has been carefully emptied out into a dish of water. Notice also that the yolk is suspended by two twisted cords of egg-white, one at either end, so that when the egg is turned the embryo always lies on top of the yolk—next to the warm body of the 'sitting' bird. The embryo, therefore, has a supply of food, water, and air, and it only needs suitable warmth in order to develop. The parent 'sits' on the eggs and keeps them warm. Cell-division and development have already begun by the time that the egg is laid, and if the egg is kept warm (103°–105° F. or about 40° C.) development continues, until after 21 days the fully-developed chick escapes from the egg-shell. If possible, examine the developing embryo in eggs that have been kept under a hen for 1, 1½, 2, 3, and 4 days.

Most birds show much *parental care* of their offspring, both before and after hatching. The embryo is given a good supply of food and water inside the egg and is protected by the hard shell. The eggs are usually laid in a carefully prepared nest, concealed from enemies, and they are kept warm until they hatch. The young birds are then fed by their parents until they can look after themselves. Many young birds remain with their parents for some time after hatching, thus learning by imitating their parents and profiting by their experience.

REPRODUCTION IN MAMMALS

Parental care reaches its highest development in *mammals*. The female parent retains the embryo inside her body throughout its development and it is fully formed at birth. After birth the young

ones are fed with milk, sheltered, and defended until they can look after themselves.

In the male mammal, e.g. cavy, rabbit, and Man, the testes begin their development in the body-cavity, close to the kidneys,

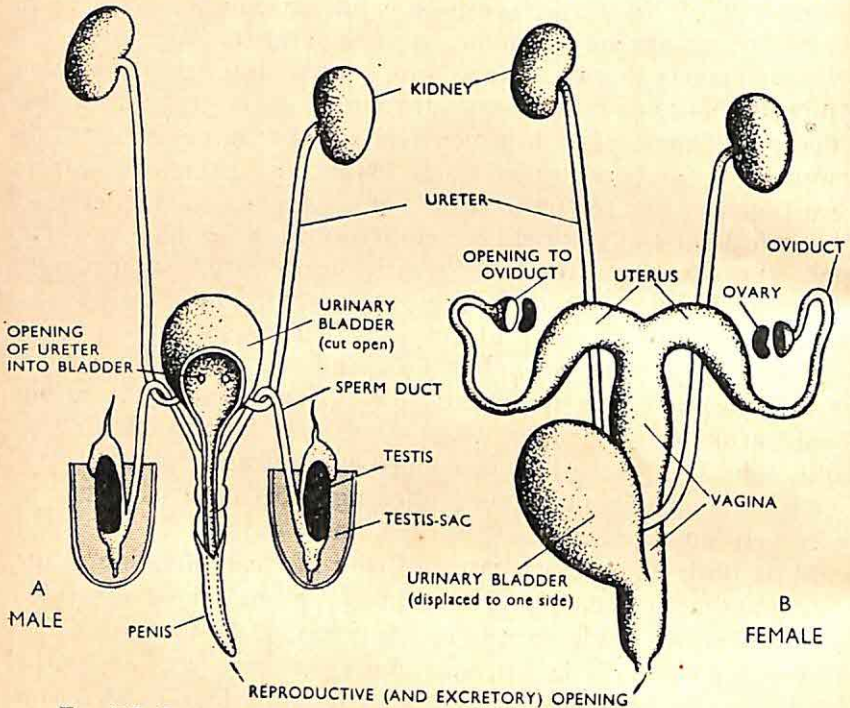


FIG. 117. Reproductive (and excretory) organs of rabbit (after Hatfield).

but before birth they pass backwards into two sacs projecting from the posterior end of the body-cavity, on the ventral side (the *testis-sacs*) (see Fig. 117). Each testis is an oval body consisting of a large number of very fine tubes from the walls of which sperms are formed. These sperm-producing tubes lead to a long, much-coiled duct in which sperms are stored until required, when they pass along a sperm-duct leading from each testis. These two sperm-ducts join together and then open into the duct leading

from the bladder to the exterior through the *penis*, or coupling organ, by means of which sperms are transferred from the body of the male into that of the female mammal. This common excretory and reproductive duct is therefore a passage for both urine and sperms. (In some mammals, e.g. cavy, there is also a pair of sperm-sacs branching off from the sperm-ducts, in which sperms are stored.)

In female mammals there are two ovaries on the dorsal wall of the body-cavity (see Fig. 117). These ovaries are very small, white, oval bodies, and they produce egg-cells that are only about 0·01 inch in diameter. Ripe egg-cells are set free at intervals, and each egg-cell finds its way into the anterior end of the oviduct, where fertilization may take place if one of the sperms received from the male swims up the oviduct and fuses with the egg-cell. The fertilized egg-cell passes down an oviduct until it reaches a wider part of the duct (the *uterus*, or *womb*), when it attaches itself to the wall of this cavity and begins to develop. (In the cavy and the rabbit there are two such cavities, and several eggs may develop at the same time, but in Man the right and left oviducts join together to form a single cavity and usually only one egg develops at a time.)

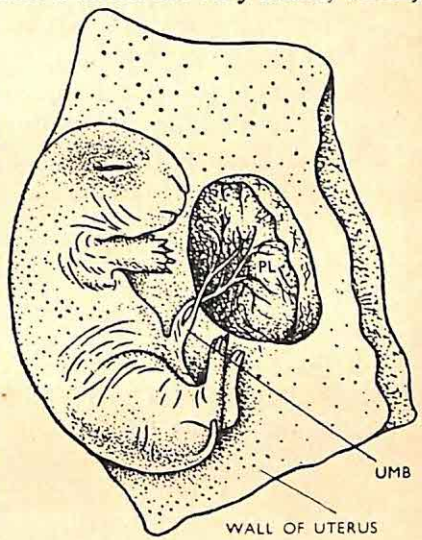


FIG. 118. Rabbit embryo before birth. PL. Disc of flesh ('afterbirth') covered with villi growing into wall of uterus. UMB. 'Birth-cord' with blood-vessels connecting embryo and disc PL (after Dakin).

The period of development inside the female parent's body (the *gestation period*†) varies in different mammals, e.g. it is about 3 weeks in rats and cavies, 9 months in Man, and 13–20 months in elephants. The developing embryo receives food and oxygen from

the mother's blood-stream, and gets rid of waste material in the same way, the capillary blood-vessels of mother and embryo being in close contact (although not actually continuous—the embryo having its own independent blood circulation). When the young mammal is fully developed, the muscular walls of the womb contract, forcing out the young animal.

Although a newly-born mammal can get its own supply of oxygen through its lungs, digest its own food, and get rid of its own waste products, most newly-born mammals are very helpless and need a great deal of parental care. They are fed on milk, produced by the milk-glands of the female parent, and this milk supplies a complete diet that requires very little digestion to prepare it for absorption and assimilation. In the highest mammal—Man—there is a long period of comparative rest, during early childhood, when parental care and protection reach their highest pitch and when the young one is trained and educated before being left to look after itself. The human body is not fully developed until about the age of twenty years, and under civilized conditions the age at which human beings begin to 'earn their own living' tends to become later and later owing to the length of time required for training and education for skilled and specialized occupations.

CELL-DIVISION

In order to understand the processes of growth and reproduction more fully, it is necessary to know something about how the body-cells multiply in plants and animals. During ordinary *growth* a single cell with its nucleus divides into two similar parts and these two halves separate and form two new cells. The division of the nucleus, however, is very complicated, and we shall only concern ourselves with its main features. Under the high power of the microscope the nucleus is seen (Fig. 119) to become a mass of fine threads (1) that shorten and thicken (2) until they become rod-like bodies called *chromosomes*† (3). (These are very important, and *every cell of the same plant or animal has the same*

number of chromosomes, which appear at every cell-division. For example, the *chromosome-number* for Man is 48, cockroach—34, cat—36, mouse—40, lily—24.) Each chromosome then splits lengthwise into two exactly similar halves (4), and the two halves

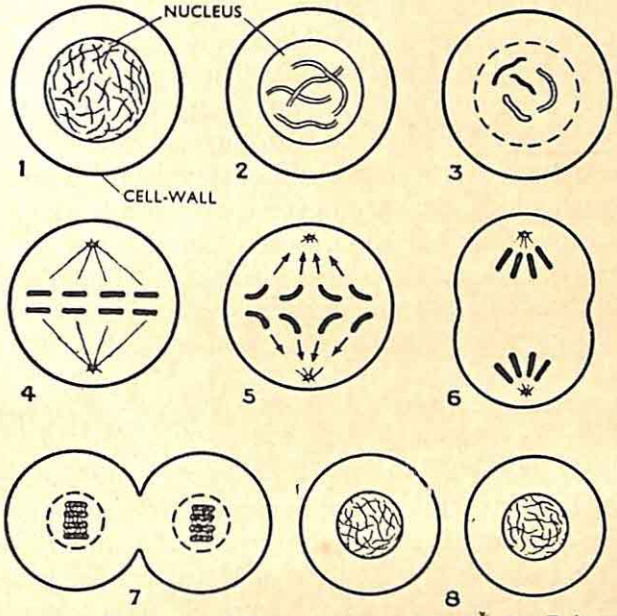


FIG. 119. Ordinary cell-division (diagrammatic—after Palmer).

are pulled apart to opposite ends of the cell (5), to form two exactly similar groups of half-chromosomes (6). Each group of chromosomes now forms a new nucleus (7) and the protoplasm of the cell divides into two halves, one half surrounding each new 'daughter-nucleus' (8).

After the two halves have separated there are two new 'daughter-cells', *each with a nucleus containing a half of every chromosome of the 'mother-cell'*. (N.B.—The terms 'mother' and 'daughter' are not used here to mean 'female'.) This whole process of cell-division usually takes an hour or two. Thus, *in ordinary cell-division, during growth and repair, every new 'daughter-cell' has exactly*

the same number and kind of chromosomes as the 'mother-cell'. These chromosomes carry the 'inheritance-units' or genes.*†* Although these *genes* are too small to be seen, even under the high power of a microscope, scientists believe in their existence because they explain the facts of inheritance (just as Dalton was forced to believe in the existence of invisible atoms because they explained the facts of chemical combination). These *genes* are the units or factors that determine the inherited characteristics of each cell. The chromosomes are collections of large numbers of genes that behave as though they were a chain of tiny particles arranged in a definite order along the chromosome, like beads along a string. During ordinary cell-division, when each chromosome splits lengthwise, each 'daughter-cell' gets an exactly similar set of genes, or 'inheritance-units'. Thus, in *asexual reproduction*, which takes place by ordinary cell-division, each new individual has the same number and kind of genes as its parent and therefore inherits the same characteristics.

CELL-DIVISION OF SEX-CELLS

In the higher plants and animals, special sex-cells—egg-cells and sperm-cells—are set apart for the purpose of *sexual reproduction*. Fertilization takes place when the nucleus of a male sperm-cell fuses with the nucleus of a female egg-cell. Hence the fertilized egg-cell will have a *double set of chromosomes, one set from the female parent and one set from the male parent*, so that if the sex-cells had the same number of chromosomes as the ordinary body-cells, then the body-cells of the offspring from sexual reproduction would have twice as many chromosomes as the body-cells of either parent. That is, the chromosome-number would be doubled in every new generation. *This is avoided by a special type of cell-division that halves the number of chromosomes during the formation of the sex-cells* (see Fig. 120). In Man, for example, during one of the last cell-divisions just before the formation of the mature sperm-cells, the 24 pairs of chromosomes do not split lengthwise (as in ordinary cell-division) but *one whole chromosome from each*

pair goes to either end of the cell; and when cell-division is complete, each mature sperm-cell has only 24 chromosomes instead of

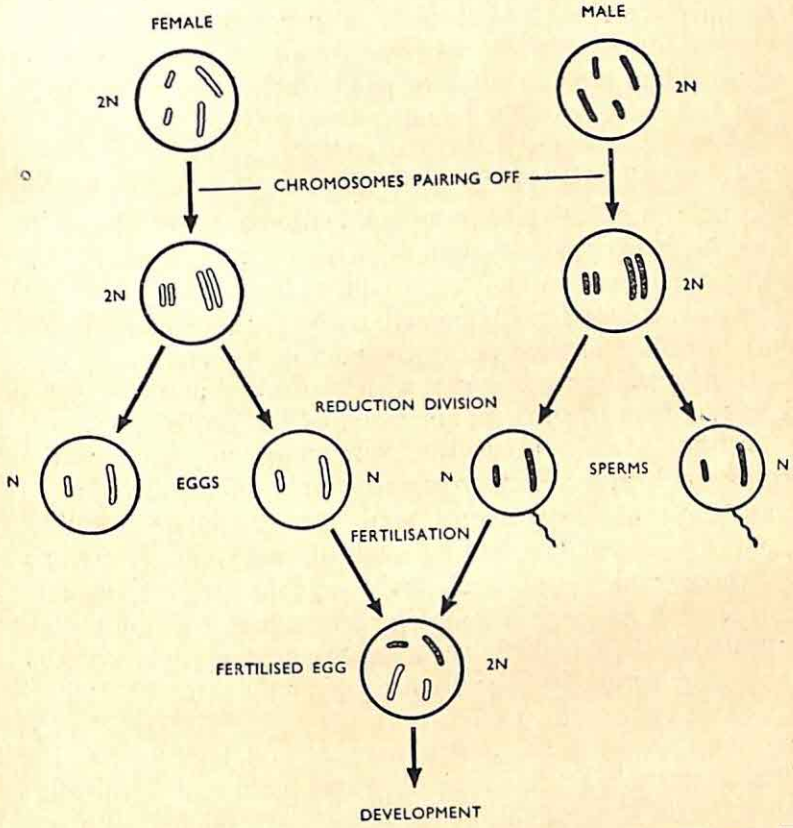


FIG. 120. Cell-division of sex-cells (final stages—diagrammatic—after Palmer). (For simplicity, this animal is supposed to have only two pairs of chromosomes per cell. The chromosomes of the male are shown in black, and those of the female in white.)

the usual 48. A similar *reduction-division* yields mature egg-cells with only 24 chromosomes. Hence, when fertilization takes place, the fertilized egg has 48 chromosomes, 24 from each parent. We see, therefore, that the life-story of any higher animal, and of any

plant that grew from a seed, begins with one single cell—the fertilized egg-cell—which carries in it all its inherited characteristics.

The fertilized egg-cell then develops into a new individual by the ordinary process of cell-division, every body-cell containing 48 chromosomes, *of which* 24 *come from each parent*. Hence the new individual inherits some of its characteristics from the male parent and some from the female parent.

INHERITANCE OF CHARACTERISTICS

We are now able to give some explanation as to (*a*) why children are like their parents (although not *exactly*), and (*b*) why children of the same two parents often differ from each other. This explanation is based partly on actual observations of the chromosomes in cells examined under the microscope and partly on the results of breeding experiments with plants and animals.

The first man to discover the essential facts of inheritance was an Austrian, Mendel, about 1865, long before anyone knew about chromosomes and the microscopic details of cell-division. His earliest breeding experiments were carried out with different varieties of garden Pea, but we shall discuss similar experiments with Maize, which is particularly convenient for breeding experiments since it has separate male and female flowers on the same plant, and if the female flower is enclosed in a paper (or Cellophane) bag before its stigmas are ripe, pollination by wind or insects can be prevented.

If the stigmas of a 'bagged' flower are pollinated by hand with pollen from the male flower on the same plant, self-fertilization is ensured. If a 'bagged' flower is pollinated by hand with pollen from a different maize plant, then cross-fertilization takes place and both parents are known.

There are different pure-bred varieties of maize, some bearing yellow grains and some bearing white grains. If a pure variety of yellow-maize is self-fertilized, all the seeds give rise to maize plants also bearing yellow grains. Similarly, self-fertilized seeds from a pure variety of white-maize all give rise to maize plants bearing

white grains. That is, these pure varieties, if self-fertilized, *breed true to the parental type.*

Now suppose that pollen is transferred by hand from the male flower of a yellow-maize plant to a 'bagged' female flower of a white-maize plant. The white-maize grains that result from this cross-fertilization, if planted, will give rise to maize plants bearing *yellow grains only.* The maize plants of this first generation of 'cross-bred' offspring, which have arisen from parents differing in the characteristic colour of their grains, are called *hybrids.*† At first sight, it would seem that the inheritance-factor for white grains (or what we may call the 'white' gene) has been lost in the first generation of hybrid offspring, but further experiments show that it is still present, although 'hidden' by the 'yellow' gene. If these hybrid yellow-maize grains are planted, and the resulting maize plants of the next generation are then self-fertilized, the maize-cobs that are formed bear *both yellow and white grains* in the proportion of 75 per cent. yellow to 25 per cent. white. Now, if all these maize-grains are planted and the resulting plants of yet another generation are again self-fertilized, *all the white-maize plants bear only white cobs* (A), *one-third of the yellow-maize plants bear only yellow cobs* (B), *while the remaining two-thirds of the yellow-maize plants bear mixed cobs with* 25 *per cent. white grains and* 75 *per cent. yellow grains.* If all the seeds from this generation are planted and the flowers are again self-fertilized, the white-maize plants from the unmixed white cobs (A) continue to 'breed true', i.e. they bear only white grains: the yellow-maize plants from the unmixed yellow cobs (B) also 'breed true', bearing only yellow grains; while the self-fertilized maize plants from the mixed cobs behave as follows:—all the white-maize plants bear pure white cobs; one-third of the yellow-maize plants bear pure yellow cobs; while the remaining two-thirds of the yellow-maize plants bear mixed cobs (25 per cent. white; 75 per cent. yellow) as before.

Figs. 121 (*i*) and (*ii*) on p. 258 will make all these results much clearer.

The explanation of this is simple. Every cell of a pure-bred

yellow-maize plant has 2 *yellow genes*, and every cell of a pure-bred white-maize plant has 2 *white genes*.

(i)

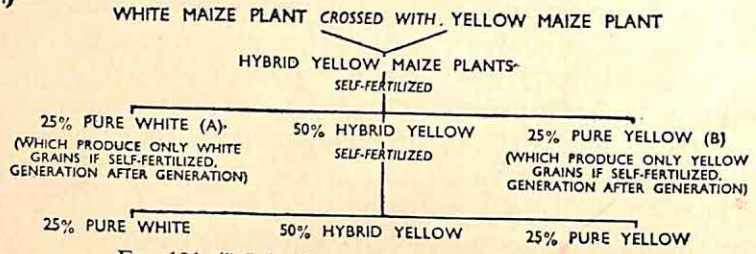


FIG. 121. (i) Inheritance of colour in maize grains.

Diagram (*i*) can now be reconstructed in terms of genes as shown below in (*ii*), representing a yellow gene by 'Y' and a white gene by 'w'.

(ii)

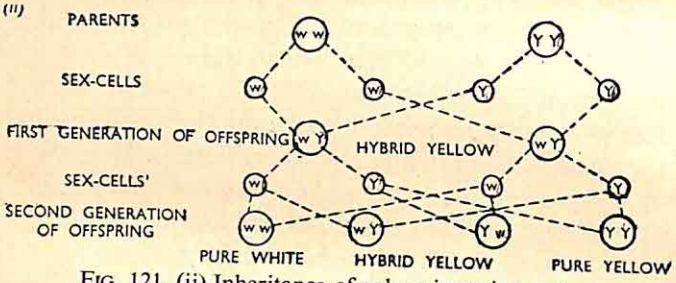


FIG. 121. (ii) Inheritance of colour in maize grains.

The colour-character that appears in the first hybrid generation (yellow in this case) is called the *dominant gene* (Y), while the character that it 'hides' (white in this case) is called the *recessive*† *gene* (w). It is very important to notice, however, that the recessive gene is merely 'hidden'—it is still present in every cell of every hybrid offspring and can always reappear as a 'pure recessive' in a later generation if another similar recessive gene is supplied during fertilization.

It is now clear why, in *sexual reproduction*, the offspring is not exactly like either parent. It gets half its genes from one parent and half from the other (and some of these genes are dominant and some are recessive) so that *sexual reproduction provides an*

opportunity for getting new varieties of plants and animals. A new variety that is well fitted for the struggle for existence will have a better chance of surviving and reproducing itself than a less well fitted new variety, so that *evolution* can take place through *natural selection.*

These principles underlying the inheritance of characteristics are applied by Man to produce new and more useful varieties of crop plants and domestic animals. Improved varieties of wheat, sugar-cane, potatoes, etc., have been obtained by crossing two widely different varieties of the same plant and then raising self-fertilized generations from this first hybrid generation. In this way thousands of different recombinations of genes are obtained from which new varieties with desirable new characteristics are produced by *artificial selection.*

(See *General Science Workbook*, Book Two, pp. 85–86, for learning exercises on this chapter.)

Section III

MECHANICS

Chapter X

MOTION AND FORCE

The word '*motion*' gives the idea of *change of position*, and we can classify most objects around us as either *moving* or *at rest*. But the terms *motion* and *rest* are only *relative*. Your school, for example, is at rest only in relation to the ground on which it stands, but the Earth's surface is not at rest since the Earth [illegible]ing around its own axis and is also moving around the Su[illegible]ll ordinary purposes, however, we describe position in rel[illegible]he Earth, and if a *body* (i.e. any material thing or *ma*[illegible]anging its position relative to the Earth, then we say th[illegible] *rest*.

SPEED AND VELOCITY

In order to describe the motion of a body completely, it is necessary to state the *direction* in which it moves and also the *speed* with which it moves. *Speed* is defined as *the distance travelled in unit time*. The term *velocity*† is used when the distance travelled in unit time is *in one particular direction*, e.g. in describing the motion of a car around a circular race-track it is correct to state its speed but not its velocity, since the *direction* of its motion is always changing.

Speed may be either *uniform* or *variable*, e.g. if a train travels exactly 88 ft. in every second, it is moving with a *uniform speed* of 88 ft. per second, i.e. 60 miles an hour. Such perfectly regular speeds are, however, most unusual in everyday life; for example, the speed of a railway train is sometimes greater and sometimes less according to the slopes and curves of the railway-track, and

the train is then said to move with *variable speed.* If a train, in spite of its variable speed, still travels 60 miles in one hour (although it has not travelled 88 ft. in each individual second) we say that the *average speed* has been 60 miles an hour. Put into symbols:

$$\text{Velocity} = \frac{\text{distance travelled}}{\text{time taken}} \quad \text{or} \quad v = \frac{s}{t};$$

i.e. distance travelled = velocity × time taken, or $s = vt$;

i.e. time taken $= \dfrac{\text{distance travelled}}{\text{velocity}}$ or $t = \dfrac{s}{v}$.

Fig. 122 shows the *velocity-time graph* of a train moving with a *uniform velocity* of 60 m.p.h. Notice that the shaded *area* gives *the distance travelled.*

THE COMPOSITION OF VELOCITIES

It is possible for a body to be moving in two different directions at the same time, e.g. when a man walks from one side of a moving ship to the other side. If the ship is travelling north with a velocity of 20 ft. per second and the man walks east at 5 ft. per second, it is clear that his actual path, relative to the Earth, is somewhere between north and east. The principle of *the Parallelogram of Velocities* states that: *If a body has two component* velocities represented by two neighbouring sides of a parallelogram, then the resultant velocity is represented by the diagonal* of the parallelogram.*

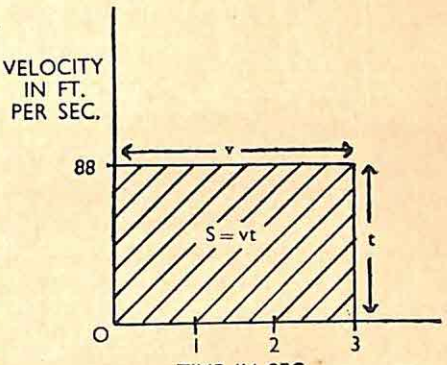


FIG. 122. Uniform velocity.

Similarly, *if a body has two component velocities represented by two sides of a triangle taken in order, then its resultant velocity is represented by the third side.*

A velocity is completely described if we can give its *rate* (e.g. in feet per second, or miles per hour) and its *direction*. Hence we can represent a velocity completely by a straight line, drawn to scale, and bearing an arrowhead. In Fig. 123, *OA* represents the velocity of the ship and *OB* the man's velocity. The *resultant velocity* is *OC*, the diagonal of the parallelogram (which in this case is a rectangle).

Work out some number problems graphically (see *General Science Workbook*, Book Two, pp. 101–4).

ACCELERATION

In everyday life, steady speeds or uniform velocities are very rare, and the velocity of most moving bodies changes from time to time. When its velocity changes, a body is said to have an *acceleration*.† This acceleration is positive if the speed is increasing and negative if it is decreasing. (A body is also said to be accelerating if the *direction* of its velocity is changing, e.g. a body moving in a circular path has an acceleration towards the centre of the circle.) *Acceleration is rate of change of velocity*. For example, if a train starting out of a station has a velocity of 2 ft. per second after 1 second, a velocity of 4 ft. per second after 2 seconds, a velocity of 6 ft. per second after 3 seconds, and a velocity of 8 ft. per second after 4 seconds, then its velocity is increasing by 2 ft. per second *in every second*; i.e. it has *an acceleration of two feet per second per second.*

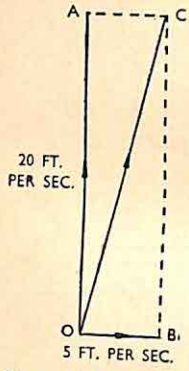


FIG. 123. Parallelogram of velocities.

Fig. 124 shows the *velocity-time graph* of the train. (The shaded *area* gives the *distance travelled.*)

Expressed in symbols: if a body starts from rest and acquires a velocity of v ft. per second in t seconds, then its acceleration

$$= \frac{v}{t} \textit{ ft. per second per second,}$$

i.e. *acceleration* = *change of velocity* ÷ *time.*

Suppose a body starts from rest and acquires a velocity of v ft. per second in t seconds, after travelling a distance of s ft.

The *acceleration* $a = \frac{v}{t}$ ft. per second per second (1*a*)

or $v = at$ (velocity = acceleration × time).

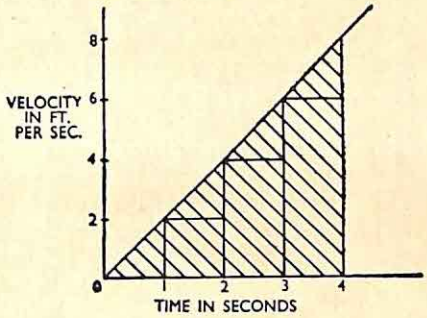


FIG. 124. Uniform acceleration.

The *average velocity* is $\frac{s}{t}$ ft. per second, and since the body started from rest, the *final velocity* v is *twice the average velocity.*

That is, $v = \frac{2s}{t}$ or $s = \frac{1}{2}tv$. (2*a*)

Hence, for a body starting from rest, we have

$$a = \frac{v}{t} \quad \text{or} \quad v = at, \quad \text{and} \quad v = \frac{2s}{t}.$$

Substituting $v = at$ in $v = \frac{2s}{t}$ we have

$$at = \frac{2s}{t} \quad \text{or} \quad s = \tfrac{1}{2}at^2. \tag{3a}$$

That is,

distance travelled = half the acceleration × the square of the time.

It is sometimes useful to express the velocity v in terms of the acceleration a and the distance travelled s.

From (1a), $\quad v = at, \quad \therefore \quad v^2 = a^2t^2.$

From (3a), $\quad s = \frac{1}{2}at^2, \quad \therefore \quad t^2 = \dfrac{2s}{a}.$

Substituting $\dfrac{2s}{a}$ for t^2 in a^2t^2, we have

$$v^2 = \frac{a^2 2s}{a}; \quad \therefore \quad v^2 = 2as. \qquad (4a)$$

Fig. 125 shows the *velocity-time graph* for a body starting from

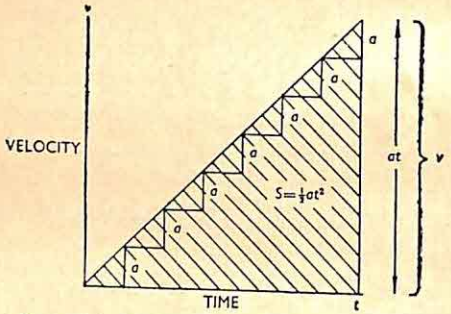


FIG. 125. Uniform acceleration from rest.

rest and moving with a uniform acceleration of a ft. per second per second.

Use these four equations of motion

$$(1a)\ v = at$$
$$(2a)\ s = \tfrac{1}{2}tv$$
$$(3a)\ s = \tfrac{1}{2}at^2$$
$$(4a)\ v^2 = 2as$$

to work out numerical problems on the motion of a body that starts from rest and travels with uniform acceleration.

TO SHOW THAT THE DISTANCE TRAVELLED IS PROPORTIONAL TO THE SQUARE OF THE TIME

Mark off a long, grooved piece of wood into four equal sections and raise one end on a wooden block as shown in Fig. 126. Start a steel ball rolling from the top of the slope and, *as it passes the first mark*, start another ball from the third mark. It is found that

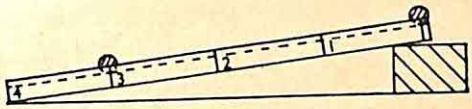


FIG. 126. One of Galileo's experiments.

the two balls both arrive at the end of the groove together. Since the time taken by the first ball to travel through the top section equals the time taken by the second ball to travel through the bottom section, the first ball will have travelled through all four sections in twice the time it took to travel through the top section. *Thus the distance it travelled is proportional to the square of the time it took* (see equation (3*a*) on p. 264).

THE EQUATIONS OF MOTION FOR A BODY THAT DOES NOT START FROM REST

Now let us consider the case of a body starting with a velocity of u ft. per second, moving for a time t seconds, travelling for a distance of s ft., and acquiring a final velocity of v ft. per second.

The *change of velocity* is $(v-u)$ ft. per second, and this takes place in t seconds.

Therefore, the acceleration

$$a = \frac{v-u}{t} \quad \text{or} \quad v-u = at, \quad \text{or} \quad v = u+at. \qquad (1b)$$

Now the *average velocity* is *half the sum of the first and the final velocities*, i.e.

$$\frac{s}{t} = \tfrac{1}{2}(v+u), \quad \text{or} \quad s = \tfrac{1}{2}t(v+u), \quad \text{or} \quad v+u = \frac{2s}{t}. \qquad (2b)$$

substituting $v = u+at$ from (1*b*), we have

$$s = \tfrac{1}{2}t(u+at+u).$$

Therefore $$s = ut+\tfrac{1}{2}at^2. \qquad (3b)$$

It is sometimes useful to express the change of velocity in terms of the acceleration a and of the distance travelled s.

Taking $v+u = \dfrac{2s}{t}$, and multiplying by $v-u = at$ we have

$$v^2-u^2 = 2as. \qquad (4b)$$

When the body starts from rest, $u = 0$, and the above equations become:

$$(1a)\quad a = \frac{v}{t} \quad \text{or} \quad v = at, \qquad (2a)\quad s = \tfrac{1}{2}tv, \qquad (3a)\quad s = \tfrac{1}{2}at^2,$$

$$(4a)\quad v^2 = 2as.$$

Fig. 127 is the *velocity-time graph* of a body starting with a velocity of u ft. per second and travelling for t seconds with a uniform acceleration of a ft. per second per second, and acquiring a final velocity of v ft. per second.

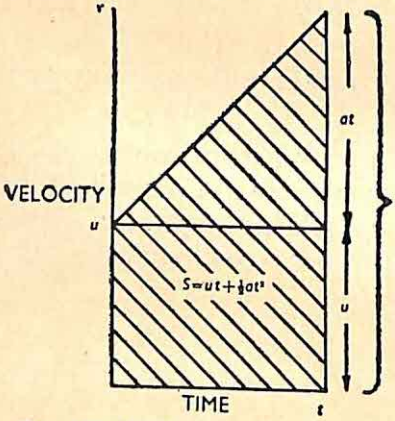


FIG. 127. Uniform acceleration of a body that does not start from rest.

Use equations (1*b*), (2*b*), (3*b*), and (4*b*) to solve some numerical examples.

FALLING BODIES

Falling bodies provide a good example of *uniform acceleration*. About 1590, Galileo (an Italian who was the leading scientist of his time) carried out experiments by dropping bodies of different sizes and materials from the top of a high tower. He concluded that *all bodies would fall at exactly the same rate if it were not for the resistance of the air*. After the invention of the *air-pump*, Sir

Isaac Newton (who was born in 1642, the year in which Galileo died) tested Galileo's conclusion by his *coin and feather experiment*. Newton enclosed a gold coin and a small feather in a long glass tube and pumped out the air. On inverting the tube he found that the 'heavy' coin and the 'light' feather fell side by side, showing that, *in a vacuum, all freely falling bodies, starting from rest, acquire exactly the same velocity after falling for the same time.*

A simple way of illustrating this principle is to place a disc of thin paper on a large coin so that the paper does not project beyond the edge of the coin. If the coin is held in a horizontal position and then released, the paper falls with it, because the coin overcomes the resistance of the air in front of the paper.

THE ACCELERATION DUE TO GRAVITY

If freely falling bodies, starting from rest, all acquire the same velocity after travelling for the same time, their acceleration must be the same. *This constant acceleration of all falling bodies is called the acceleration due to gravity*, and it is usually represented by the symbol g.

The value of g varies very slightly in different parts of the world. Thus in London it is 32·19 ft. per second per second; in New York it is 32·16 ft. per second per second; in Calcutta it is 32·11 ft. per second per second; and in Singapore $g = 32{\cdot}09$ ft. per second per second. We shall take $g = 32$ *ft. per second per second* as being accurate enough for practical purposes. On the *centimetre/gram/second* (C.G.S.) system, $g = 981$ *cm. per second per second.* (As you may have a clearer practical idea of speeds in miles per hour, it may be useful to remember that g is an acceleration of about 22 m.p.h. per second.) From the equation $s = \frac{1}{2}at^2$ you will see that all bodies fall 16 ft. earthwards during the first second after they begin to fall from rest.

We have seen that the *weight* of a body is not exactly the same everywhere on the Earth's surface, being greater at the Poles than at the Equator. It is clear that the variation in the value of g and the variation in the *weight* of a body are due to the same cause.

What is it that keeps moving bodies in motion? And what stops them? The old idea of motion, as taught by the early Greek scientists, was that 'force is necessary to maintain motion'. They thought that a body must be continuously pushed or pulled

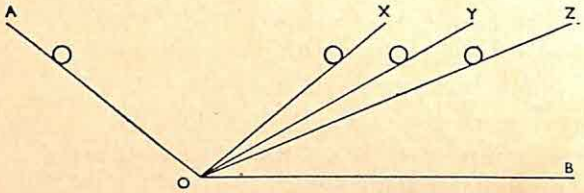


FIG. 128. Another of Galileo's experiments.

to keep it in motion, and that as soon as the force producing motion was removed the body would stop. Galileo showed that this was incorrect, and he discovered the underlying principles, later put into formal laws by Newton. Galileo arrived at his conclusions from an experiment with two inclined boards (see Fig. 128).

When a ball was rolled down the smooth inclined plane *AO* and then up any of the planes *OX*, *OY*, *OZ*, he found that the ball climbed very nearly to the vertical height from which it had descended. As the angle, *XOB*, is reduced (e.g. to *ZOB*) the ball travels farther in climbing nearly to the original height, so Galileo argued that if the plane *OX* were lowered until it became horizontal (*OB*) then, *if there were no friction*, the ball would travel on for an indefinitely great distance. Under such conditions there is no horizontal component (see pp. 279-80) of the force of gravity acting on the ball, nor is there any opposing frictional force, and we conclude that a body in motion will continue to move with uniform velocity when no force is acting on it.

Newton's First Law of Motion states that: *A body continues in its state of rest, or of uniform motion in a straight line, unless it is acted on by a force.* Hence no non-living body at rest can move

of itself, and no non-living moving body can change its speed or direction *of itself*.

This is a fundamental property of matter, i.e. any attempt to move it when at rest, or to stop it when in motion, is resisted. This universal property is called *the inertia*† *of matter*.

It is this property of *inertia* that keeps the planets from falling on to the Sun; that makes mud and water fly off at a tangent from a revolving bicycle wheel; that makes the Earth's diameter greater at the Equator than at the Poles.

Man has never made a machine that, of itself, will continue in a state of uniform motion for ever. He never will succeed in making such a 'perpetual* motion machine', because it is impossible to get rid of friction entirely. In fact, Newton's First Law cannot be *proved experimentally* because it is impossible to get rid of friction. However, there are many pieces of evidence pointing to the truth of the law. For example: (1) *Astronomical predictions* that apply the law come true*; e.g. using the law, men can accurately forecast the position of a planet on any future date and also the date of an eclipse. (2) The more friction is diminished the more nearly does the motion of a body become uniform, e.g. when a ball is bowled in turn along a cricket pitch, a wooden floor, a sheet of ice, it rolls farthest on ice and its motion is nearly uniform. (3) If you are cycling at a high speed and your bicycle strikes an obstacle,* you will be thrown forward over the handle-bars. (4) A car skids if the brakes are suddenly applied when the road is wet or muddy. (5) A standing passenger is thrown backward when a train starts or is thrown forward when the train stops suddenly.

Can you think of other illustrations of the law?

FORCE

Several kinds of force are already familiar to you: (1) *The force of gravity*. We have found the *weight* of bodies by measuring the pull of the Earth upon them. (2) *Magnetic force*. We have seen that the N-pole of a bar magnet exerts a force of attraction (or pull) on the S-poles of other bar magnets and a force of repulsion

(or push) on their N-poles. (3) Forces brought about by the change in shape of a body, i.e. by its expansion or contraction; e.g. the *tension* (or pull) of a stretched spring or of an animal muscle; the *thrust* (or push) exerted by growing seedlings forcing their way through soil. (4) *Friction*: the force that opposes the motion of a body when it is in contact with another body.

Up to the present we have been content to describe force as a *push* or a *pull*. We know, however, what a force *does*, and it is the *effects* produced by forces that concern us in everyday life. Perhaps the most satisfactory definition of force is in terms of Newton's First Law of Motion (although this law does not tell us how to *measure* force): *Force is that which changes, or tends to change, the state of rest of a body, or its uniform motion in a straight line.* For example: (i) A force may cause a body to move from rest. (ii) A force may cause a moving body to move faster or slower (or stop it moving). (iii) A force may change the direction of a moving body. (iv) A force may set up a *strain* in an elastic body that is unable to move.

FORCES THAT PRODUCE NO MOTION

Forces do not always produce movement. Thus, when you sit down on a chair the force of gravity is pulling you downwards (as you notice when someone removes the chair before you sit down!) but your body remains at rest because the upward pressure of the chair exactly balances your weight. These two *equal and opposite forces*, which produce no motion, are said to be *in equilibrium*.†

In the same way, when an object is hung from a spring-balance it descends until the upward pull of the spring is equal to the downward pull of gravity, when it comes to rest, the two forces being balanced or *in equilibrium*. These examples also illustrate Newton's Third Law of Motion (or the Law of Reactions): *To every action there is an equal and opposite reaction.* Consider a book resting on a table. It is acted upon by gravity, and if this were the only force acting on the book it would produce motion and the book would fall vertically downwards through the table. The book

remains at rest because the table presses vertically upwards on the book with an equal and opposite force.

The following simple experiments illustrate this principle of equal and opposite reactions: (*a*) Hook together two spring-balances and pull them apart. Compare their readings. (*b*) Try to walk along a board (about 6 ft. long and 10 in. wide) resting on wooden rollers (of about 1 in. diameter) placed about 18 in. apart on a level floor. The board slides backward as fast as you walk forward. Walk backward along the board; it moves forward. (*c*) Blow up a rubber balloon and pinch its neck between finger and thumb. Throw it up into the air. The air inside the balloon rushes out backward and the balloon travels forward. This illustrates the principle of rocket- and jet-propulsion.*

THE FORCE OF GRAVITY—WEIGHT AS A FORCE

This subject was introduced in Book One, Chap. VI. Newton, after many experiments and observations, concluded that: (*a*) All objects attract each other. (*b*) The force of attraction depends upon the distance between the objects (i.e. the nearer together two objects are, the greater the pull between them). (*c*) The force of attraction depends on the *mass* or quantity of matter in the objects (i.e. massive bodies, like the Sun, pull more strongly than smaller bodies). (*d*) Objects near the Earth fall towards the centre of the Earth because they are continuously attracted by its enormous *mass*. (They accelerate as they fall.)

The force of attraction between all objects is called the *force of gravitation*. When one of the attracting bodies is the Earth, we speak of the *force of gravity*. The Earth is the largest mass that concerns our daily life and we notice its pull on other objects. This pull of the Earth on an object is called the *weight* of the object. Thus *weight is a force* and the weight of a body is the force with which the Earth and the body attract each other.

THE FORCE OF FRICTION

When you push very gently against a book resting on a table,

at first it will not move, but, as you increase the pushing force, it begins to *slide*. To *keep it moving* you have to apply *less* force than you did to *start it moving*. You are now applying a force against the opposing force of *frictional resistance*. This force of friction exists whenever one body moves in contact with another, and it acts in the opposite direction to the force that keeps one body moving. The *static*† *friction* force, which must be counterbalanced before the body begins to move, is always greater than the *kinetic*† *friction* force that must be applied to keep the body moving at a uniform speed in contact with another body.

Friction is of enormous importance to the engineer and he is always trying to reduce friction. This can be done by choosing the right kind of bearing surfaces, by making these very smooth, and by oiling them. The thin layer of oil slides with the moving surface and the friction is now between a liquid and a solid; this is always much less than between two solids.

As bodies *rolling* over one another are less subject to friction than bodies *sliding* over one another, *ball bearings* and *roller bearings* are often used instead of plain bearings. The more the friction can be reduced the less energy is wasted, for some of the energy involved when frictional forces do work is converted into useless or even dangerous heat (as when a bearing '*seizes*').

Sometimes the engineer makes use of friction, as when he designs *brakes*, usually made of materials between which the frictional forces are very great.

THE MEASUREMENT OF FORCE

The units commonly used to measure forces are the *pound weight* and *gram weight*. One pound weight is the pull of the Earth and a *mass* of one pound on each other. If a lump of iron is held on the hand we feel it pressing downwards. If the hand is taken away, the lump of iron, still attracted by the Earth, falls to the ground. If we hang the iron from a spring, the stretch of the spring measures the pull of the Earth, i.e. the weight of the iron. Hence,

the simplest method of measuring forces is by means of a *spring balance*.

The heavier the body, the greater is the stretch of the spring. The relation between the force applied to a spring and the stretch (or *extension*) it produces is stated in Hooke's Law: *Within certain limits, the stretch (or extension) is directly proportional to the stretching force (or load)*. This law can be verified by experiment.

Set up the given spring and metre rule as shown in Fig. 129. A piece of bent sheet-metal, attached to the lower end of the spring by wax, serves as a pointer for measuring changes in length. Take the reading of the pointer when there is *no load* on the scale-pan and record it below. Then place a 10 gm. weight on the scale-pan and take the new reading. Repeat this, increasing the load by 10 gm. wt. each time until the total load is . . . gm. wt. Then decrease the load, 10 gm. wt. at a time, and notice whether the extension is the same as before.

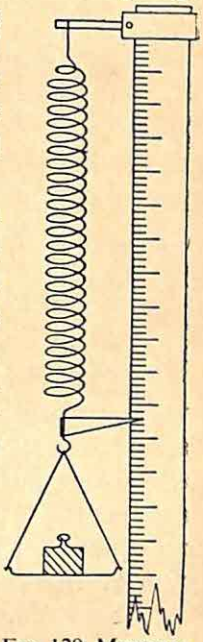


FIG. 129. Measurement of force.

Enter your results in columns:

LOAD gm. wt.	TOTAL LENGTH FROM SCALE	TOTAL EXTENSION	EXTENSION PER 10 gm. wt.
0			
10			
20			
etc.			

The spring-balance makes use of this principle that a spring stretches uniformly, and it is suitably 'scaled' to measure weights.

ELASTICITY

The spring used in the above experiment *recovered its original*

length when the stretching force was removed. When a body or material recovers its original shape or size in this way it is said to be *elastic*. But a spiral spring is elastic only up to a certain limit; it is possible to add so much load that the spring will not recover its original length when the load is removed, but remains *permanently stretched*. The maximum load that the spring can carry without becoming permanently stretched is called its *elastic limit*. Lead wire is said to have very little *elasticity* because the elastic limit is reached almost at once when it is stretched. Rubber, however, is said to possess great elasticity because it can carry a considerable load without being permanently stretched. A fuller statement of Hooke's Law, therefore, says that when an elastic body is stretched the extension is directly proportional to the load, *provided that the elastic limit is not reached.*

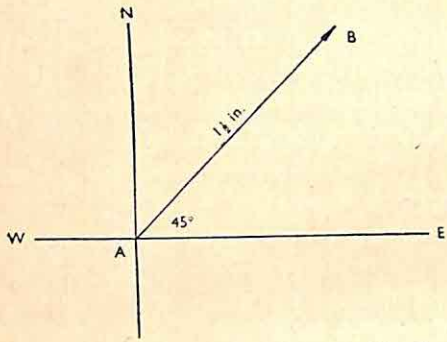


FIG. 130. A force represented graphically.

In Book Two, Chap. VII, we saw that *sound* is the result of *wave motion*. Sound waves can be produced *only in elastic materials*. Resistance to change of *shape* causes *transverse waves in solids*. Resistance to change of *size* causes *longitudinal waves* (or *compression waves*) in elastic solids, liquids, and gases.

HOW FORCES ARE REPRESENTED GRAPHICALLY

Forces have *size* (which we measure in units of weight) and *direction*. They can therefore be represented graphically by straight lines; the *length* of the line representing the *size* of the force, and the *direction* of the line representing the *direction* in which the force acts. (An arrowhead on the line shows which way the force acts along the line.) To represent a force completely, we must also know at what point of the body it acts, because our

force-line must start from a definite point. For example, in Fig. 130, the line AB represents a force of 15 lb. wt. acting at the point A in a north-easterly direction, the scale being 1 in. = 10 lb. wt.

ADDITION OF FORCES

This is the process of finding that single force which is equivalent in action to a number of forces. The single force that will produce the same effect as the joint action of several component forces is called their *resultant*.†

COMPOSITION AND RESOLUTION OF FORCES

The resultant of two forces is that single force which will produce the same effect as the joint action of the two forces. The substitution of a single force to produce the same effect as a number of forces acting together is called *the composition of forces.* The converse substitution of two or more forces that act together to produce the same effect as a single force is called *the resolution*† *of forces.*

When two forces acting *in the same direction* are applied to the same point, the *resultant force* is equal to the *sum* of the two forces. Similarly, when two forces acting in opposite directions are applied to the same point, the resultant force is equal to the *difference* between the two forces and acts in the direction of the greater force. In other words, *when a number of forces act on a body in the same straight line, their resultant force is the ALGEBRAIC SUM of the component forces.* For example, a telephone post supports two wires, one having a tension of 100 lb. wt. and the other a tension of 80 lb. wt. It is clear that if one wire runs north and the other wire south from the post, then the resultant force is a pull of 20 lb. wt. in the direction of the greater pull, i.e. north. If both wires run south from the post, then the resultant force is a pull of 180 lb. wt. in the same direction.

We must now consider forces that do not act in the same straight line, but at an angle to each other.

INCLINED FORCES ACTING AT A POINT

(i) Tie three different weights, A (= ..... lb. wt.),

B (=.....lb. wt.), C (=.....lb. wt.), to three pieces of string knotted together at O, and hang the weights A and B over two pulleys as shown in Fig. 131 (a). When the system comes

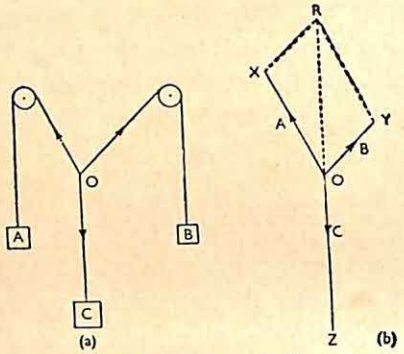


FIG. 131. Forces in equilibrium.

to rest we have, acting at the point O, three forces that are in equilibrium. It is clear that the downward force C is balancing the inclined forces A and B, and the force C is called the *equilibrant*† of A and B.

This force C must be equal and opposite to the *resultant* of A and B.

(ii) Place a sheet of paper behind the strings and mark the direction of all three strings on the paper (see Fig. 131 (b)). Remove the paper and mark off lengths OX, OY, and OZ, proportional to the weights A (.....), B (.....), and C (.....) respectively. Produce ZO to R so that $ZO = OR$. Then OR represents the *resultant* of A and B in size and direction. Join XR and YR and examine the figure $OYRX$. If your pulleys are easy-running, if your strings are light, and if you have done the experiment carefully, $OYRX$ is a *parallelogram*.

We have now a simple graphical method for finding the resultant of two inclined forces acting at a point, using the *Principle of the Parallelogram of Forces*: *If two forces are represented in size and direction by two neighbouring sides of a parallelogram, then their*

resultant is represented in size and direction by the diagonal of the parallelogram. For example, in Fig. 132, if *OA* and *OB* represent

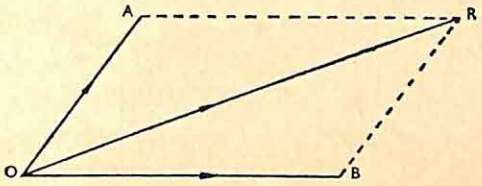


FIG. 132. Parallelogram of forces.

two forces drawn to scale, then *OR* represents their resultant drawn to the same scale.

(iii) On the same sheet of paper that you used in Expt. (ii), draw three lines parallel to the forces *OX*, *OY*, and *OZ*, in such a way as to form a triangle (see Fig. 133). Measure the sides of the triangle.

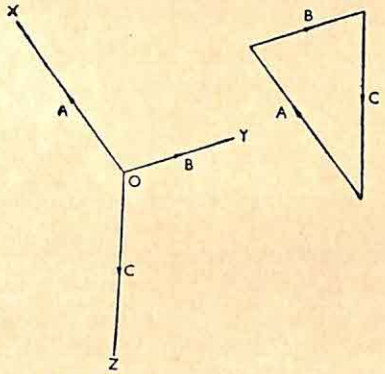


FIG. 133. Triangle of forces.

They are proportional to the forces *OX*, *OY*, and *OZ*. This result illustrates *the Principle of the Triangle of Forces*: *If three forces acting at a point are in equilibrium, they can be represented in size and direction by the three sides of a triangle taken in order.* Conversely, *if three forces acting at a point can be represented in size and direction by the three sides of a closed triangle taken in order, then they are in equilibrium.* Hence, when you try to draw a

triangle of forces and it does not close, the three forces are *not* in equilibrium.

Notice that it saves time to draw the triangle OYR instead of completing the whole parallelogram $OYRX$.

The *parallelogram of forces* and the *triangle of forces* apply to systems of three forces only. A similar principle states the conditions for equilibrium between more than three forces: *If any*

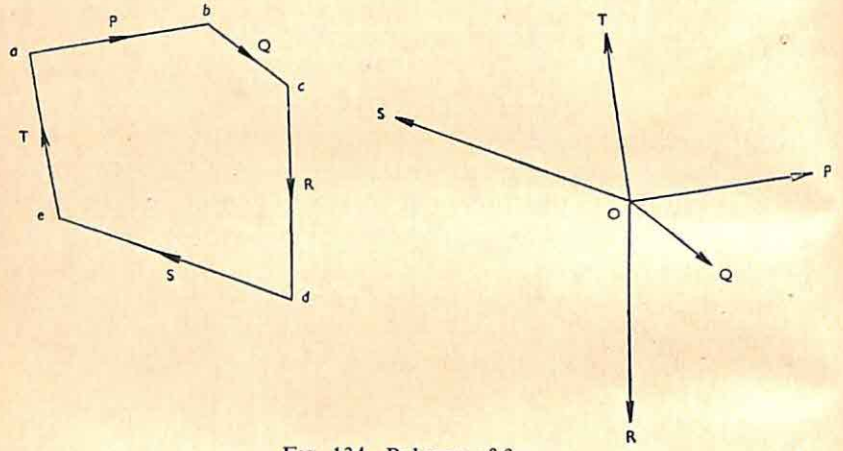


FIG. 134. Polygon of forces.

number of forces P, Q, R, S, T, etc., acting at a point are in equilibrium, they can be represented in size and direction by the sides of a closed polygon† abcde, etc., taken in order (see Fig. 134). Conversely, *if a system of forces acting at a point can be represented in size and direction by the sides of a closed polygon taken in order, then the forces are in equilibrium.* Hence, when you try to draw a polygon of forces and it does not close, the extra line required to close the figure represents the *equilibrant* of the unbalanced forces (see Fig. 135, where AE represents the *equilibrant* R).

The application of these principles is of the greatest practical importance to engineers that design structures and have to work out the forces acting on the various parts, e.g. in the framework of a steel bridge. If you are to understand these principles (as distinct

from merely learning them by heart) you must practise applying them by solving as many simple number problems as possible, by graphical methods (see *General Science Workbook*, Book Two).

THE RESOLUTION OF FORCES

We have seen how two or more forces can be replaced by a single force called their *resultant*.

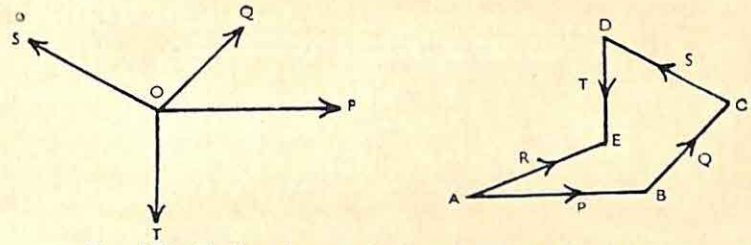


FIG. 135. Finding the equilibrant of four unbalanced forces.

It is sometimes necessary to replace a single force by two or more forces that, acting together, will produce the same effect. This is called the *resolution of forces* since one force is *resolved* or converted into two or more forces acting in different directions.

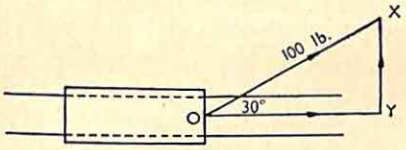


FIG. 136. Resolution of forces.

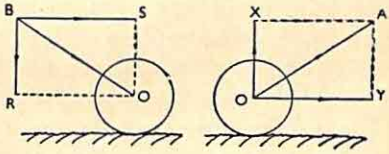


FIG. 137. Resoultion of forces

For example, in Fig. 136 a small railway coach is being pulled along the rails by a man X walking outside the rails, so that the rope OX makes an angle of 30° with the rails. If the man's pull is 100 lb. wt., find how much of this pull is doing useful work in moving the coach along the rails, and how much is being wasted in pulling the wheels at right-angles to the rails. We have to *resolve* the force of 100 lb. wt. into two other forces along OY and OX. We therefore mark off a length OX proportional to 100 lb. wt. (say

2 in.) and then draw XY perpendicular to OY. Then OY represents the useful forward pulling component (= ..... in. = ... lb. wt.) and XY represents the side-thrust wasted in pressing against one of the rails (= ..... in. = ... lb. wt.).

To take another example, why is it easier to *pull* a garden-roller on a soft piece of ground than to *push* it? See Fig. 137. Resolving the *push BO* into a horizontal component and a vertical component,

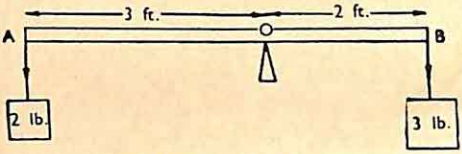


FIG. 138. Parallel forces.

the horizontal force BS tends to move the roller along the surface while the vertical force BR acts *downwards*, pushing the roller deeper into the ground. Resolving the *pull OA* into a horizontal force and a vertical force, the horizontal force OY tends to move the roller along the surface while the vertical force OX acts *upwards* and reduces the pressure of the roller on the ground; so that it moves more easily when it is pulled than when it is pushed.

Solve some simple number problems graphically for yourself.

PARALLEL FORCES

We have now considered (*a*) forces acting in the same straight line and (*b*) forces inclined at an angle to each other. The remaining possibility is (*c*) forces acting parallel to each other.

If two forces act in parallel directions we cannot find their resultant by using the parallelogram or triangle of forces. For example, if masses of 2 lb. and 3 lb. are hung from each end of a bar AB, 5 ft. long, as shown in Fig. 138, there will be two parallel forces acting on the bar. If we put a knife-edge under the bar, the force at B tends to move in a clockwise direction and the force at A tends to move in an anti-clockwise direction; but we can find one point where the bar will balance (at O) and the resultant of the

two downward forces must act in a vertical line passing through this point O. The point O is called *the centre of the two parallel forces*, and it is in such a position that *each downward force multiplied by its distance from O gives the same product*. This product

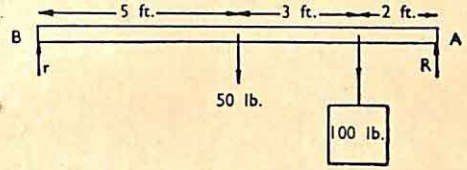


FIG. 139. Parallel forces.

is called the *leverage* (or the '*moment*'†) of the forces about the point O and, *when the bar is in equilibrium, the clockwise leverage at B is equal to the anti-clockwise leverage at A*. This is *the Principle of Leverage* (or '*moments*') and it also applies to any number of parallel forces acting on a bar.

In Fig. 138 the *leverage* of the force at A is 2 (lb. wt.) $\times$ 3 (ft.) = 6, while the *leverage* of the force at B is 3 (lb. wt.) $\times$ 2 (ft.) = 6. In this case the *resultant* is a force of 5 lb. wt. acting vertically downwards at O, which is 3 ft. from A and 2 ft. from B (neglecting the weight of the bar AB).

The Principle of Leverage (or 'moments') enables us to calculate the forces acting on a loaded beam or bridge. For example, in Fig. 139, AB is a uniform beam, 10 ft. long and of mass 50 lb., resting on two supports at A and B, and a mass of 100 lb. is hung 2 ft. from A. The downward forces acting on the two supports A and B are equal to the upward reactions R and r at A and B.

As the beam is in equilibrium, the upward forces R and r must balance the downward forces, i.e. $R+r = 50+100 = 150$ lb. wt., and the clockwise leverage about either A or B must be equal to the anti-clockwise leverage. Consider the leverage about A:

$$\begin{aligned} 10r &= (100\times 2)+(50\times 5) \\ &= 200+250 \\ &= 450. \end{aligned}$$

Therefore $r = 45$ lb. wt. But $R+r = 150$ lb. wt.

Therefore $R = 105$ lb. wt.

We can check this by considering the leverage about B:

$$10R = (50\times 5)+(100\times 8)$$
$$= 250+800$$
$$= 1050.$$

Therefore $R = 105$ lb. wt.

EQUILIBRIUM

Equilibrium is a state of balance when a number of forces acting on a body produce no motion, i.e. when the body is at rest under the action of several forces.

CENTRE OF GRAVITY (OR CENTRE OF MASS)

Substances are composed of vast numbers of small particles (molecules). Each particle is attracted towards the centre of the Earth. Thus we have a great number of parallel forces. We may find the resultant of any two of these forces and, having found this single force, we can, with a third parallel force, find the resultant that will be equal to all three forces combined—and so on with a fourth, etc. Finally, we shall have one single force (W) that is the resultant of the weights of all the particles (see Fig. 140). This resultant (W), equal to the weight of the body, passes through a point C, whose position remains fixed in the body however this is placed. This fixed point is called the *centre of gravity* of the body.

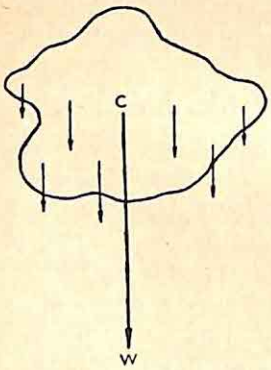


FIG. 140. Centre of gravity.

The centre of gravity is that point in the body through which its total weight (considered as a single force) always acts, no matter what position the body may take up.

This total weight, or resultant (W), can be counter-balanced by a single force acting vertically upward and from beneath W. Thus the centre of gravity is directly above the point where the body

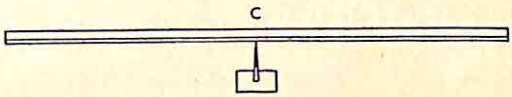


FIG. 141. Centre of gravity and point of balance.

balances, e.g. a ruler balanced on a sharp point (see Fig. 141) has its centre of gravity at C.

If a body is pivoted at its centre of gravity it will remain in any position in which it is placed. When a body is hung from a thread, it always comes to rest with its centre of gravity vertically below the point of support.

(i) Find the centre of gravity of an irregularly shaped piece of cardboard as follows: Hang the sheet by a pin, e.g. at A (see Fig. 142). The sheet will hang so that its centre of gravity is vertically below the point of support. If a plumb-line is hung from the pin, the centre of gravity will lie somewhere along this line, which should be drawn on the cardboard. Now hang up the sheet by a pin through B and draw the vertical line through B in the same way. Then the centre of gravity lies somewhere along these two lines. It must therefore be at their point of intersection C.

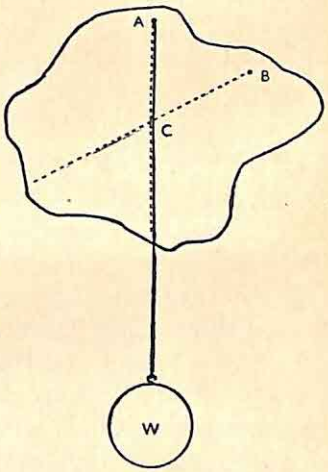


FIG. 142. Finding the centre of gravity.

(ii) Trace a map of your country on a sheet of cardboard and then cut round the outline with scissors. Find the best place for a 'central' broadcasting station (from the point of view of distance only).

(iii) Find the centre of gravity of (*a*) a retort stand, (*b*) a tripod stand.

STABILITY

The *stability*† of a body (i.e. whether it is easy to upset or not) depends upon the position of its centre of gravity in relation to its point of support.

A body is said to be in *stable equilibrium* if it returns to its

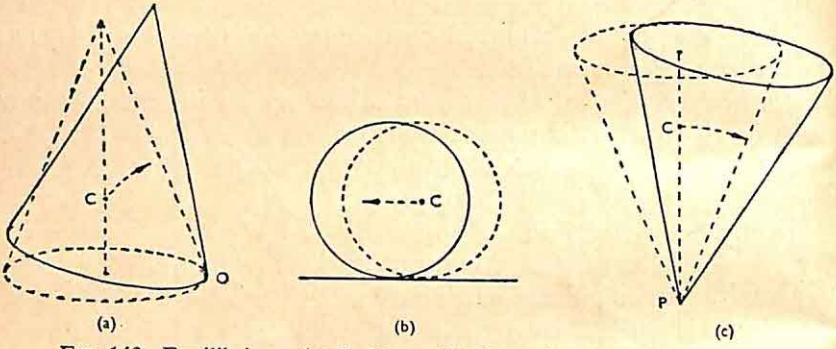


FIG. 143. Equilibrium. (a) Stable equilibrium. (b) Neutral equilibrium. (c) Unstable equilibrium.

original position after slight displacement, e.g. a pendulum, a brick resting on its larger face, a cone or a filter-funnel standing on its base. For a body to be in *stable equilibrium, its centre of gravity must be as low as it can get.* See Fig. 143 (a).

A body is said to be in *neutral equilibrium* when, after a slight displacement, it remains in whatever position it has been put, e.g. a sphere or cylinder lying on a horizontal surface, a cone or a filter-funnel lying on its side. A body is in neutral equilibrium if its centre of gravity is neither raised nor lowered when it is displaced slightly. See Fig. 143 (b).

A body is said to be in *unstable equilibrium* when, after a slight displacement, it moves still farther from its original position, e.g. a cone balanced on its point. A body is in unstable equilibrium when a slight displacement *lowers its centre of gravity*, i.e. its

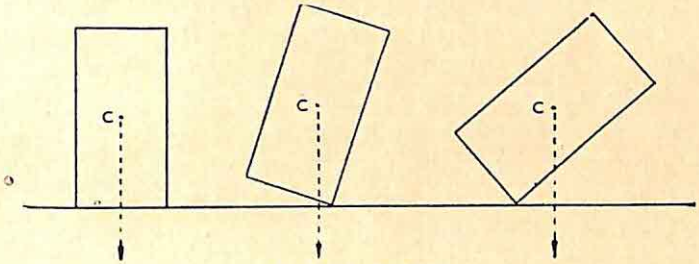


FIG. 144. Stability.

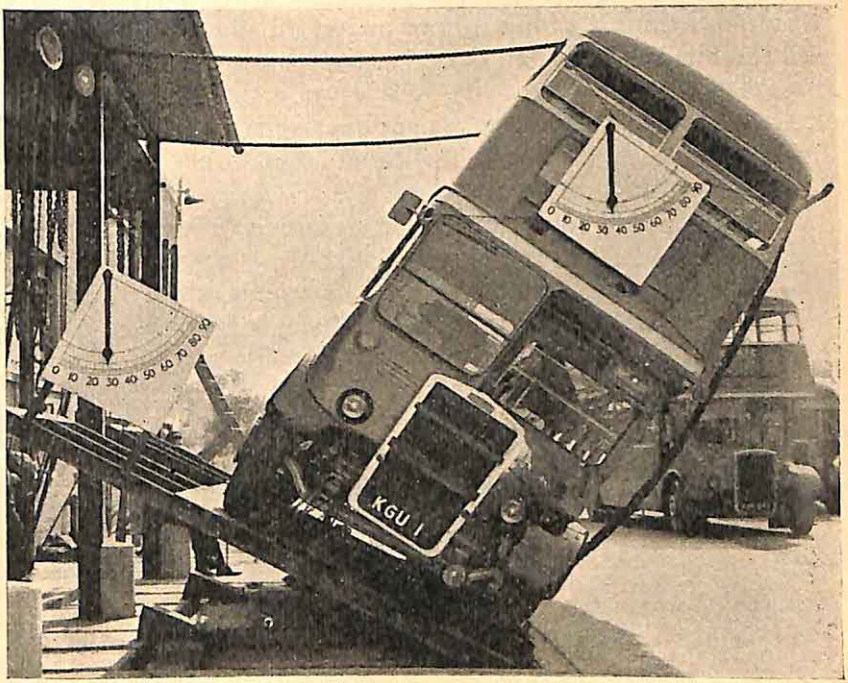


FIG. 145. Bus being tested for stability.

centre of gravity was originally vertically above the point of support. See Fig. 143 (c).

For any object resting on a support to be in *stable* equilibrium, *the vertical line drawn through its centre of gravity must fall within its base* (see Fig. 144). Stand with your heels against a wall and try to pick up a book from the floor in front of you without bending your knees and without moving your feet. This will enable you to verify the truth of the above principle.

Any article that must not upset easily, e.g. a motor-car, is therefore built with the engine and all the heavier parts *as low as possible*, so as to produce a low centre of gravity. Similarly, a car with wide wheel-tracks is less likely to overturn than one with narrower wheel-tracks. Motor-buses are usually tested for stability by putting them on a hinged floor and gradually raising one side until the bus begins to fall over, i.e. until the vertical line drawn through the centre of gravity falls outside the wheel-track. A modern 'double-decked' passenger-bus when fully loaded will stand on a sideways slope of 30° without falling over (see Fig. 145).

In this photograph, why does the *clinometer*† (or angle-measurer) attached to the bus show a reading different from the clinometer attached to the hinged floor? In such a stability test, bags of sand are placed on the seats to represent the weight of the bus-crew and passengers. In practice, a bus never uses a road with such a steep side-slope, but the big *margin of safety* provides for the effect of cornering at speed and of strong side-winds.

(See *General Science Workbook*, Book Two, pp. 87–108, for numerical problems and learning exercises on this chapter.)

CHAPTER XI

ENERGY

Throughout this Course we have frequently distinguished between *matter* and *energy*. We have a fairly clear idea of what we mean by matter—it is something that we can weigh, something that 'occupies space'. We know that *light* and *heat* and *sound* are not forms of matter; we call them forms of energy.

When anything is able to *do work*, it is said to possess *energy*, i.e. *energy is the capacity for doing work* and it is measured in the same units as work. We must therefore learn what scientists mean by 'work' and how they measure it.

WORK

Most people think of 'work' as anything that involves 'labour' or muscular *effort*. But the scientific meaning of 'work' is more precise:* *When a force moves its point of application, it is said to do work*. From the scientific point of view, therefore, work is always associated with *moving* something, e.g. lifting a weight upwards (against gravity) or pushing something sideways (against friction), and no work is accomplished until a force *overcomes* a resistance. (In much the same way, workmen are often paid, not for their *efforts*, but for *the effect produced*, or *work done*.)

The *quantity of work done* is measured by *the product of the force and the distance moved*, i.e. *work* = *force* × *distance*. For practical and engineering purposes, the *unit of work* is the *foot-pound*, and this can be defined as *the work done in lifting a mass of one pound through a vertical height of one foot*. Thus, if you lift a 10 lb. weight from the floor and put it on a table 3 ft. high, you have done 30 foot-pounds of work.

Notice carefully that, in the scientific sense, no work is done until the force succeeds in *moving* the body on which it acts. Thus,

if you push against a heavy motor-car you are exerting a force on the car and may, indeed, become fatigued by pushing, but you do not perform any *work* until the car *moves*. It is important to remember this difference between the use of the term *work* in the everyday sense and in its strict scientific meaning.

POWER

When we measure work in foot-pounds, we take no account of the *time* taken to do the work. For many purposes, it is important to know the *rate of working*, and this is called *power*. *Power* (or *rate of doing work*) can be expressed in foot-pounds per second, or as foot-pounds per minute. Engineers, however, use the *horse-power*† as their practical unit.

This *British unit of power* was introduced by James Watt (1736–1819), the inventor of the modern steam-engine, who considered that an average horse could do 33,000 foot-pounds of work per minute (or 550 foot-pounds per second) over an average working day. *An engine is developing one horse-power* (1 *h.p.*) *when it is doing work at the rate of* 33,000 *foot-pounds per minute.*

Measure your own horse-power as follows: Run up a long flight of steps as fast as you can while someone with a stop-watch takes the time as you reach the bottom step and as you pass the top step. In this period, you have raised your own weight through the vertical height of the steps and can now calculate your *rate of working*:

Weight = ...lb. wt. Number of steps = ...ft.

Height of one step = ...in.

Therefore, vertical height of .....steps = ...ft.

Time taken = ...seconds.

$$\text{Horse-power developed} = \frac{\text{wt. in lb.} \times \text{height in ft.}}{\text{time in seconds} \times 550} = \frac{\times}{\times\ 550} = \ldots . \text{ h.p.}$$

This will be nearly your maximum rate of working, which you can maintain only for a few seconds. Over a longer period, your rate of working would be only about 0·1 h.p.

The *metric and electrical unit of power is* called the *watt*† (after James Watt) and the power of electric motors is usually expressed in *Kilowatts*† (1 Kilowatt = 1,000 watts). A horse-power = 746 watts (or nearly $\frac{3}{4}$ Kilowatt), i.e. 1 Kilowatt = $1\frac{1}{3}$ horse-power (nearly). We shall discuss later the relation between the watt and other electrical units (see pp. 367–8).

The rate at which an engine or motor works can be measured as shown in Fig. 146. A belt or cord is placed over a pulley on the engine, one end of the belt carrying a mass of, say, 35 lb., and the other end being attached to a spring-balance, which, in turn, is fastened to the ground or to the bench. When the engine is at rest, the spring-balance will read 35 lb. wt.; but when the engine is running, the friction between the pulley and the belt will tend to raise the weight, and the spring-balance will show a smaller force, say 5 lb. wt. If the speed of the pulley is 1,100 revolutions per minute, and its circumference is 1 ft., we can find the rate at which the engine is doing work.

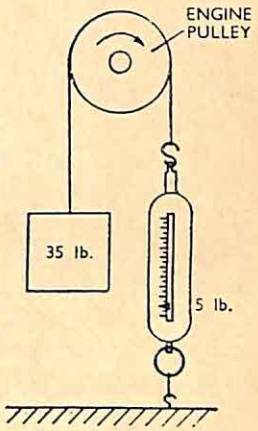


FIG. 146. Measurement of power.

In one complete revolution of the pulley, the engine does 35—5 = 30 foot-pounds of work against friction, hence the work done per minute is 30 × 1,100 = 33,000 foot-pounds per minute, i.e. the engine develops 1 h.p.

The power of an engine measured in this way is called its *brake horse-power* (b.h.p.). Fig. 147 shows one form of practical brake-apparatus for this purpose.

ENERGY

In the earlier parts of this Course we have often referred to different forms of *energy*; in fact, at least one-third of the whole Course is devoted to *physics*, which is largely *the study of the effects of different forms of energy on matter*. In biology lessons, too, we

have seen that the fundamental living processes of *photo-synthesis* and *respiration* involve transformations of energy.

From very early times, Man has made use of the energy supplied by animals to help him to do work. But every year he is learning

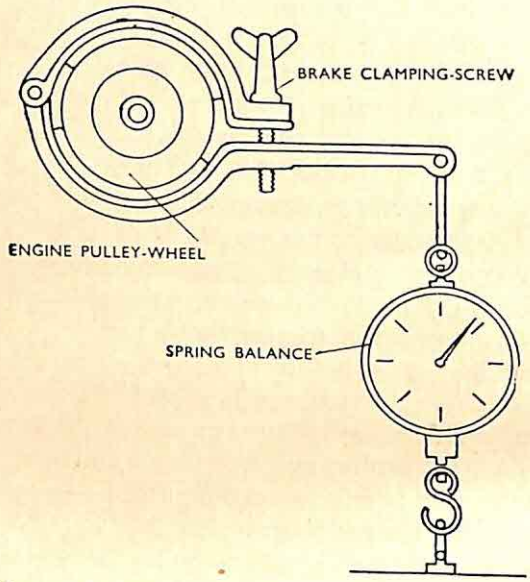


FIG. 147. Brake-apparatus for measuring horse-power (after Harrison).

to make more use of the energy in Nature in different ways. He uses running water to turn water-wheels and turbines that drive dynamos and provide him with electrical energy: he still uses wind energy in sailing-ships and windmills; he uses the energy locked up in coal and oil to drive his steam-engines and his internal combustion engines; he uses the energy stored in plant and animal foods in his own growth and movement and in keeping himself warm.

POTENTIAL ENERGY

A huge storage reservoir high up in the mountains looks quiet

and harmless, but it possesses a great store of energy because, if released, the water rushes down towards sea-level and on the way *it does work*. This work may be 'useful', as when we use the water to drive a water-wheel, or it may be harmful, as when the work is done in uprooting trees and destroying buildings. We say that the water in the raised reservoir possesses *potential energy*.†

A body possesses potential energy because of its *position* or its *condition*. All forms of energy, except that due to *motion*, are classed as potential energy or *stored* energy. For example, a 10 lb. weight on a table 3 ft. above the floor has potential energy because, by using a suitable machine, it can be made to do 30 foot-pounds of work as it falls to the floor. It owes its energy to its *position*, like the water in a raised reservoir. On the other hand, coiled clock springs, explosives like dynamite, fuels like coal and food, all contain potential energy because of their *condition*. The spring may *unwind* and keep the clockwork moving (doing work) for several days; the explosive may *decompose* violently and shatter a rock or propel a bullet; the coal and the food may be decomposed by *oxidation* and release enough energy to drive a steam-engine or to enable an animal to move.

KINETIC ENERGY

A moving body of any kind possesses energy. We have defined *energy* as *the capacity for doing work*, and we have seen that it is measured in units of work, and also that work always involves movement. Hence the only form of energy that can be measured *directly* in work-units is *energy of motion* (which is sometimes called *mechanical energy* or *kinetic energy*†). *The kinetic energy of a moving body is the amount of work it will do before it is brought to rest.* For example, the kinetic energy of moving air (wind) can be made to do useful work in driving sailing-ships and in turning windmills; while moving water (streams and rivers) can be made to turn water-wheels and turbines.

The moving parts of all our engines possess kinetic energy and

do work; so do growing roots and stems, and moving animals of all kinds.

When a small clock or watch is wound up, work is done in compressing the elastic spring, and the spring gains potential energy equal to the work done in winding it up. While the clock is working, this potential energy, stored up in the compressed spring, is gradually transformed into the kinetic energy of moving parts of the clock.

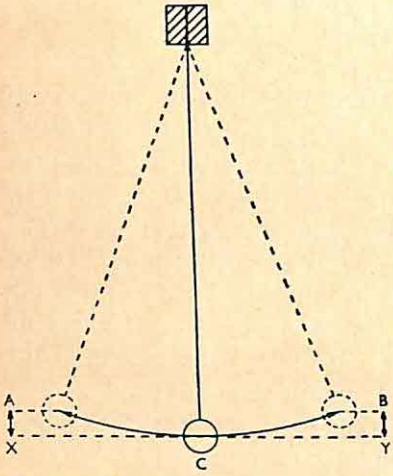


FIG. 148. Potential and kinetic energy.

The *pendulum* shows clearly this transformation from potential energy to kinetic energy and back again. As shown in Fig. 148, when the pendulum 'bob' is at rest at *C*, it possesses neither kinetic energy nor potential energy since it has no motion and is as low as it can be. If we pull it aside to position *A*, we do work in raising it a vertical height *XA*, and this work is stored up in the bob as potential energy. If we release the bob, it swings along the arc *ACB*. When the bob reaches the point *C*, at the bottom of its swing, the whole of its potential energy is converted into kinetic energy, which is then reconverted into potential energy once more by the time the bob reaches the point *B*. At *A* and *B*, therefore, the energy of the swinging pendulum bob is entirely *potential*. As it passes *C*, its energy is entirely *kinetic*. At points between *A* and *C* and between *C* and *B*, its energy is partly kinetic energy and partly potential energy, the sum of the two being equal to the original potential energy when at *A*.

WATER-WHEELS AND TURBINES

In the same way, the water at the top of a waterfall possesses *potential energy* owing to its position. If a water-wheel is placed in

the path of the falling water, the water falls into the buckets at the top of the wheel and the potential energy of the water is transformed into the kinetic energy of the moving wheel before the buckets allow the water to escape at the bottom of the wheel. Water-wheels (which have now been largely displaced by water-turbines

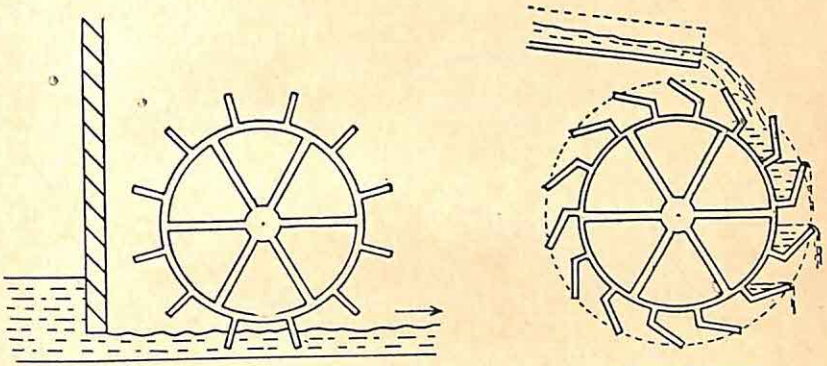


FIG. 149. Water-wheels.

and Pelton wheels† for large-scale purposes) are of two main types: *undershot wheels* and *overshot wheels*, as shown in Fig. 149. The *undershot wheel*, used in flat country, is placed with its lower part in the running water, and some of the *kinetic energy* of the moving water (usually about 25 per cent.) is transferred to the wheel. In the *overshot wheel*, used in hilly country, the running water falls over the top of the wheel so that use is made of both the *kinetic energy of the running water* and the *potential energy of the falling water*. This is a much more efficient arrangement than the undershot wheel, and over 80 per cent. of the potential energy of the water can be converted by an overshot wheel into useful work.

In hydro-electric power-stations these simple water-wheels have been displaced by *water-turbines* and by *Pelton wheels*, which are more efficient. Thus, some of the 100,000 h.p. turbines at Niagara Falls have an efficiency of over 90 per cent. Water-turbines are used where the 'head' of water is only a few hundred feet, and they

run at a relatively low speed. When the 'head' of water is over 500 ft., a Pelton wheel is used (see Fig. 150). The Pelton wheel is a special type of undershot wheel in which the high-pressure water is delivered from a jet and strikes the cup-shaped buckets at a very high speed. The buckets are shaped so that they return the water

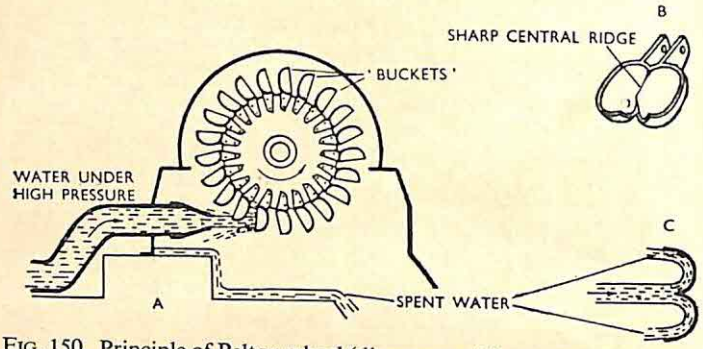


FIG. 150. Principle of Pelton wheel (diagrammatic). A. Vertical section showing wheel carrying ring of spoon-shaped 'buckets'. B. Single 'bucket' removed from wheel. C. Section through 'bucket' showing how jet of water is split and turned through nearly 180°, thus transferring its kinetic energy to the revolving wheel.

in the direction opposite from that in which it strikes the wheel; thus, when the wheel is revolving, if the speed of the buckets is half the speed of the jet of water, most of its kinetic energy (over 85 per cent.) is given up to the wheel (Fig. 150). The largest Pelton wheels used in hydro-electric power-stations develop about 30,000 h.p.

TRANSFORMATION OF ENERGY

Energy can exist in several different forms, and each kind of energy can be transformed into any other kind. Some common forms of energy are *heat energy*, *light energy*, *sound energy*, *electrical energy*, and *chemical energy*, and these can all be converted into each other. For example, green plants absorb *light energy* and store it up in carbohydrates as *chemical energy*. When plant remains (e.g. wood or coal) are burned under a boiler, this chemical

energy is converted first into *heat energy* and then into the *potential energy* of high-pressure steam. If this high-pressure steam works a *steam-turbine*, its potential energy is converted into *kinetic energy* as it passes through the turbine, and this kinetic energy is given up to the moving parts of the turbine. If a dynamo is coupled to the turbine, this kinetic energy (or mechanical energy) is converted into *electrical energy*. This electrical energy can be converted either into *light* and *heat* (in an electric bulb), or into *kinetic energy* (in an electric motor), or into *sound* (in an electric bell or electric gramophone), or into *chemical energy* (in a lead accumulator or during electrolysis).

It is possible, theoretically, to change all of these different kinds of energy into each other, although in practice some changes are easier than others. Notice particularly that electrical energy, even after travelling for long distances through wires, can be changed easily and directly into all the other forms of energy: this is one reason why electricity is so useful in everyday life.

Transformations of energy are of great importance in living things. The *light energy* of the Sun is transformed in green leaves into the chemical energy of carbohydrate foods. Later this energy is changed again, by respiration; some is released as *heat energy*, some becomes *kinetic energy* of moving parts, some is reconverted into *chemical energy* during the synthesis of new chemical substances that build up the organism.

CONSERVATION OF ENERGY

Careful experiments have shown that *in all these transformations of energy, no energy is ever lost or destroyed but is only changed into other forms of energy*. This principle is called the *Law of Conservation of Energy*, i.e. *energy cannot be destroyed*, and the total amount of energy in the Universe* always remains the same; it only changes from one form into another and finally into *heat*. This *Law of Conservation of Energy* is the basis of quantitative *physics*, just as the *Law of Conservation of Matter* is the basis of quantitative *chemistry*.

Although every transformation of energy yields an equivalent amount of other forms of energy, in actual practice it is seldom possible to *use* all the energy or reverse the transformation exactly, because of unavoidable losses. For example, no practical heat-engine converts *all* the heat of the working substance into mechanical energy; the used steam or exhaust gases still contain some heat energy that merely warms up the atmosphere and which cannot be recovered. This is *wasted energy* but it is *not* destroyed (although it is no longer available for doing *useful work*).

Even the most carefully-made machines always waste some energy in *friction* and *noise*, which finally yield heat energy at too low a temperature to be of practical use. Animals, also, though more efficient with their energy conversions than most machines, do not often use more than about 25 per cent. of their available energy in doing useful work—the rest is lost as heat.

When the driver of a motor-car wishes to stop quickly, he has to change kinetic energy into some form of potential energy, and he does this by applying the brakes. The brake-drums get hot owing to the conversion of kinetic energy into heat. In theory it should be possible to start the car again by reconverting this heat into kinetic energy, but in practice the heat escapes into the atmosphere and a fresh supply of kinetic energy is obtained by burning petrol in the cylinders of the engine and transforming the heat energy into energy of motion.

When a raised weight falls freely, all its potential energy is converted into kinetic energy by the time it reaches the ground. On striking the ground, this kinetic energy is not destroyed but is changed into *heat*. In the same way, the water in a pool at the bottom of a waterfall possesses less potential energy than before it fell, but its temperature is slightly higher because this potential energy has been changed into heat.

All the energy we use today comes from the Sun.[1] The energy

[1] In a few places (e.g. the Bay of Fundy in Canada) the rise and fall of sea-tides is used as a source of tidal power. The original source of energy in this case is the gravitational force of the Moon.

of sunlight is fixed in food, in oil, and in coal. Radiant heat from the Sun is transformed into the energy of moving air (winds); it is used in the evaporation of water that later falls again as rain on the hills, where it may supply the kinetic energy of waterfalls. Scientists are now beginning to use the energy locked up in atoms; but even the atoms of the elements came from the Sun when the Earth was first formed.

(See *General Science Workbook*, Book Two, pp. 87–108, for numerical problems and learning exercises on this chapter.)

CHAPTER XII

MACHINES

Man, in his primitive state, had to depend upon his own muscles and limbs like all the other animals. It has taken about 20,000 years for Man to reach his present state of civilization and his dominant position in the world. He owes much of this advance to learning

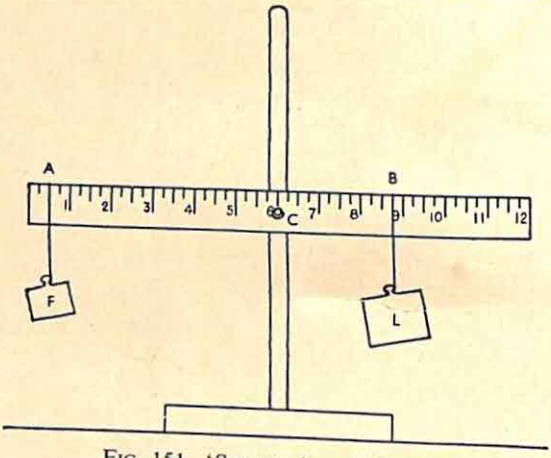


FIG. 151. 'See-saw lever' (Class I).

gradually how to use *tools* and *machines*. With the aid of these he can produce and control forces that would be far beyond his power if he used his body alone.

A *machine* is an appliance by which forces can be made to produce other forces of different *magnitude** and *direction*. The animal body itself contains machines (levers) whereby forces exerted by the muscles move the limbs, the head, and the trunk. However, we commonly use the word machine to mean anything by which a man can apply a force more conveniently and more usefully than with his unaided limbs. Sometimes a machine *magnifies* the effect

of the force we apply, e.g. a bicycle is a machine that enables your leg muscles to move your body farther and faster and more easily than by walking. Man uses some machines to transform energy, e.g. a dynamo is a machine for transforming kinetic energy into

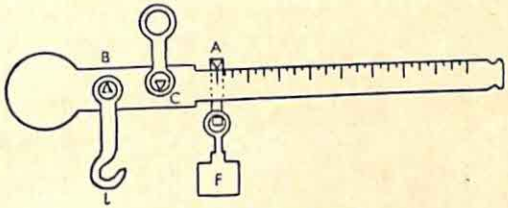


FIG. 152. Steelyard.

electrical energy; a steam-engine is a machine for transforming heat energy into kinetic energy. Man also uses machines to *transfer* energy from one place to another. In this chapter we shall consider what scientists call the *simple machines*: (i) *levers*, (ii) *pulleys*,† (iii) *wheels and axles*, (iv) *inclined planes*, (v) *wedges*, (vi) *screws*.

LEVERS

The apparatus shown in Fig. 151 is a simple *lever*, i.e. *a rod free to turn about some fixed point C called the pivot*. It is clear that such a lever can be used as a *machine*, the *force* (or *effort*) acting at *A* and overcoming a *load* (or *resistance*) acting at *B*. The pivot need not always be placed between the effort and the load: in fact, there are three possible arrangements of the *pivot*, the *force*, and the *load*, giving rise to three classes of levers.

Class I. 'See-saw'† levers. In levers of the *first class the pivot is between the force and the load*, as shown in Fig. 151, where the *force F* acts downwards at *A* and the *load L* acts downwards at *B*, on the opposite side of the *pivot C*. Examples of such levers are (i) a see-saw, (ii) the beam of a balance, (iii) a steelyard† used for weighing heavy articles (see Fig. 152), (iv) a pair of scissors or wire-cutters, (v) a crowbar† when used in the ordinary way for raising a heavy stone (see Fig. 153).

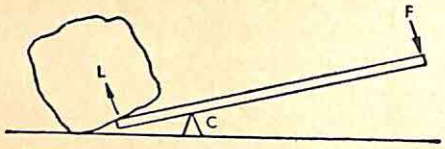


FIG. 153. One way of using a crow-bar.

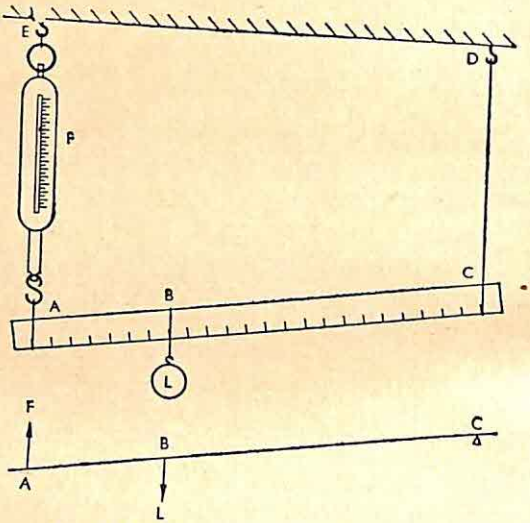


FIG. 154. 'Wheelbarrow lever' (Class II).

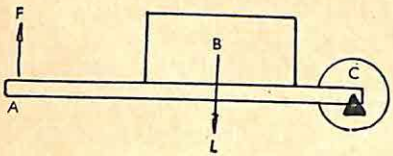


FIG. 155. Principle of wheelbarrow.

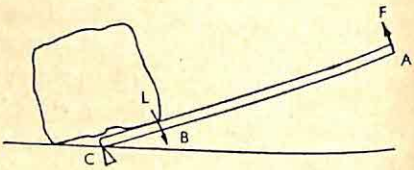


FIG. 156. Another way of using a crowbar.

Class II. 'Wheelbarrow'† levers. In levers of the *second class the load is between the force and the pivot*, as shown in Fig. 154, where the *force F* acts upwards at *A* and the *load L* acts downwards at *B*, between the force F and the *pivot C*. Examples of the second

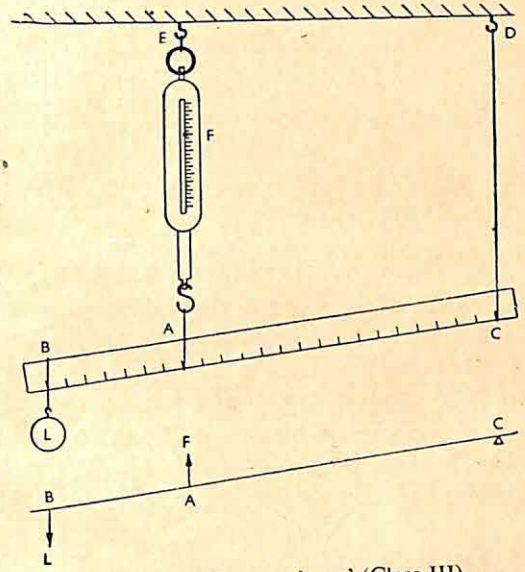


FIG. 157. 'Forearm lever' (Class III).

class of levers are (i) a wheelbarrow (see Fig. 155), (ii) the human foot and ankle (see Fig. 159), (iii) a bottle-opener for 'crown corks', (iv) a crowbar when used as in Fig. 156.

Class III. 'Forearm' levers. In levers of the *third class the force acts between the load and the pivot*, as shown in Fig. 157, where the *load L* acts downwards at *B* and the *force F* acts upwards at *A*, between the load *L* and the *pivot C*. Examples of the third class of levers are (i) the human forearm when lifting a weight with the hand (see Fig. 159), (ii) a pair of forceps, (iii) a steam-engine safety valve (see Fig. 158).

Set up levers of each class as shown in Figs. 151, 154, and 157. Use large weights so that the weight of the lever-beam itself is

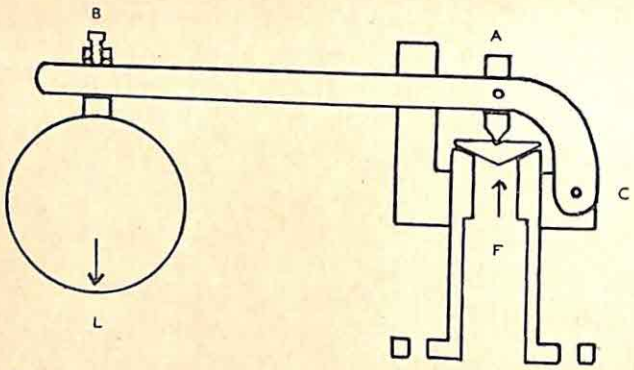


FIG. 158. Steam-engine safety-valve (diagrammatic section).

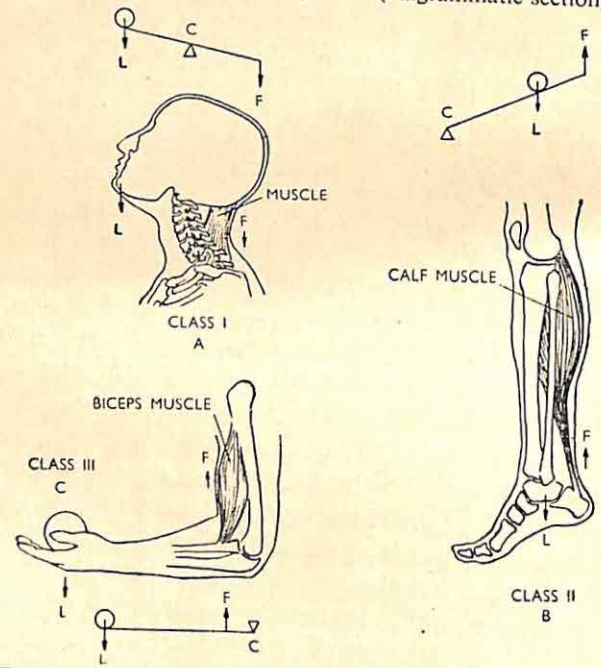


FIG. 159. The three types of animal levers. A. First type of lever and its action at joint between skull and backbone. B. Second type of lever and its action at ankle joint. C. Third type of lever and its action at elbow joint.

small in proportion to the force and the load. Enter your results below:

Class of Lever	I	II	III
Force (F)			
Distance of force from pivot (*x*) . . .			
Load (*L*)			
Distance of load from pivot (*y*) . . .			
Leverage of force (*Fx*) ('*force-arm*') . .			
Leverage of load (*Ly*) ('*load-arm*') . .			

If you have worked carefully, you should find that the last two quantities *Fx* and *Ly* are equal, i.e. *force-arm* = *load-arm*. This is called *the principle of the lever*.

Notice that levers of the first and second classes can give a mechanical advantage greater than unity, i.e. a small force applied as the *effort* can move a larger *load*. Levers of the third class, however, never give a mechanical advantage since the *effort* is always greater than the *load*, but they give *convenience* (e.g. in forceps) and *speed* (e.g. any movement of the biceps muscle is greatly magnified in the corresponding movement of the hand).

Use *the Principle of Levers* to solve some number problems (see *General Science Workbook*, Book Two, pp. 94–98).

PULLEYS AND LIFTING-TACKLE

If you want to raise a heavy object to the top of a building there are several ways of doing it. One way is to carry the object up the stairs or up a ladder, but this involves raising your own body as well as the object, thus 'wasting work'. Thus, if a man weighing 160 lb. carries a load of 40 lb. wt. up a ladder to a height of 50 ft. he does 10,000 foot-pounds of work, 2,000 foot-pounds of which represent 'useful work' and the remaining 8,000 foot-pounds represent 'wasted work'.

Another way is to stand at the top of the building and pull up the object with a rope. Nearly all this work is 'useful work', but

it is usually more *convenient* to pass the rope over a *pulley* fixed to the top of the building and pull *down* on the rope while standing on the ground. To lift the load of 40 lb. wt. through 1 ft. you would still have to do 40 foot-pounds of work by pulling the rope downward

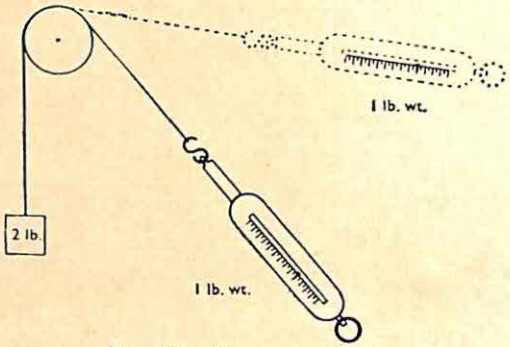


FIG. 160. Single fixed pulley.

with a force of 40 lb. wt., i.e. *a single fixed pulley merely changes the direction of the pull*, making it more *convenient*, but it gives *no mechanical advantage* (and it does not multiply speed).

Pass a string over a single fixed pulley as shown in Fig. 160. Hang a 2 lb. weight on one end and fasten a spring-balance to the other. Hold the spring-balance in different positions so that the string attached to it makes different angles with the vertical. What is the reading of the spring-balance in each case? ..... lb. wt.

In everyday life it is often necessary to lift a big weight (or *load*) by applying a small force (or *effort*) so as to get a *mechanical advantage*,† which is defined as:

$$\text{Mechanical advantage} = \frac{\text{load}}{\text{force}}.$$

The *mechanical advantage* (or *force ratio*) of any machine, therefore, is *the ratio of the load raised to the force required to keep the machine moving at constant speed.*

If we turn the last arrangement (Fig. 160) upside-down (as shown in Fig. 161) we can increase the effect of our pull. *Half* the

2 lb. weight is now supported by one string and *half* by the other string. The *force F* required to hold the *load L* in place is therefore only *half* the load = 1 lb. wt. Verify this by experiment, measuring the force *F* with a spring-balance as shown in Fig. 161.

Such an arrangement of pulleys is called a *tackle*,† and in this

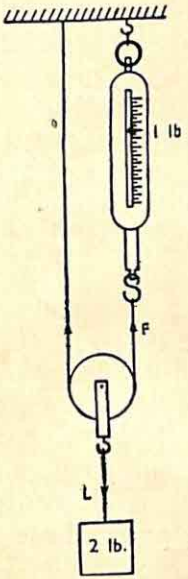


FIG. 161. Single movable pulley.

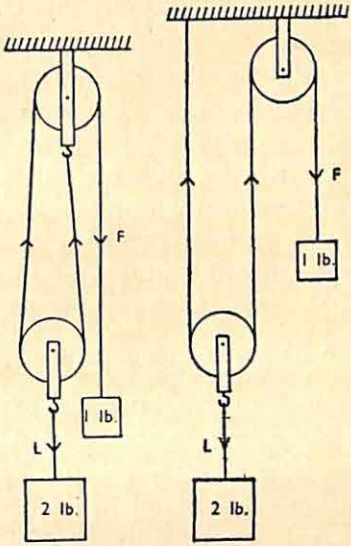


FIG. 162. Simple lifting tackle.

case we can lift a load twice as great as the force applied to the string, i.e. this tackle gives us a *mechanical advantage* of 2. In practice the arrangement shown in Fig. 161 is inconvenient because it requires an upward pull. We have seen, however, that a single fixed pulley will change the direction of the pull, hence the arrangements shown in Fig. 162 are more convenient.

Fit up one of the arrangements shown in Fig. 162.

(i) First balance the movable pulley with hanging weights. Then hang a 2 lb. weight from the movable pulley and find what force is required at *F* to balance this load. .....lb. wt.

(You can do this either by adding more weights at F or by using a spring-balance.)

What is the mechanical advantage? .....

(ii) Measure the distance that the load L rises when the force F descends 1 ft. .....ft. The ratio between these two distances is called the *velocity ratio* of the machine.

$$\text{Velocity ratio} = \frac{\text{distance moved by force}}{\text{distance moved by load}} = \frac{\quad}{\quad} = \ldots\ldots \text{ in}$$

this case. Notice that in this case (when the movable pulley is counter-balanced*) the *mechanical advantage is practically the same as the velocity ratio*, but this is only true of an 'ideal' lifting tackle in which the movable pulleys have no weight and in which all the pulleys turn without friction.

(iii) Now remove the hanging weights you used to counter-balance the movable pulley, and find the force F required to raise a load of 2 lb. wt. using a spring-balance. Force = ...lb. wt.

$$\text{Therefore, actual mechanical advantage} = \frac{\text{force}}{\text{load}} = \frac{\quad}{\quad} = \ldots\ldots$$

The velocity ratio, however, is unchanged. In this case, therefore (and *in all practical machines*), *the mechanical advantage is less than the velocity ratio* because some work is always wasted in overcoming *friction* and in lifting moving parts whose weight is not counted as part of the 'useful load'.

THE PRINCIPLE OF WORK

In Expt. (ii) above (where friction was too small to affect our rough measurements) we see that the product *load* × *distance moved by load* is the same as the product *force* × *distance moved by force*, i.e. *the WORK done on the load BY the machine = the WORK done by the force ON the machine.*

We found the same relation in the hydraulic press (Book Two, Chap. IV) where the force acting on the small piston × the distance through which it moved = the force acting on the large piston × the distance through which it moved, i.e. *the work put into the machine is equal to the work got out of the machine* (if there is no friction).

This important *Principle of Work* was first stated by Sir Isaac Newton in 1687, and it applies to all machines. In other words, *a machine cannot do more work than is done ON it*. Hence, although a machine enables us to increase the *force* available, it does *not* enable us to do *more work*.

Most of the *simple machines* multiply the force applied to the machine. In such machines, the *time* is increased; so what we gain in force we lose in time or speed. A bicycle, however, is an example of a machine that is used to *gain speed* and we increase our speed by increasing our effort. The bones of the human arm are levers forming a machine that we use to gain speed of movement of the hand.

THE EFFICIENCY OF A MACHINE

In all actual, practical machines, some energy is wasted (as heat) in overcoming friction, i.e. *the work put into machine = the work got out of machine + the work done in overcoming friction*: so that *less 'useful' work* is got out of the machine than is put in. This fraction (usually expressed as a percentage) is called the *efficiency*† of the machine.

$$\text{Efficiency} = \frac{\text{'useful' work got out of machine}}{\text{total work put into machine}} = \frac{\text{output}}{\text{input}}$$

$$= \frac{\text{load} \times \text{distance load moves}}{\text{force} \times \text{distance force moves}}$$

$$= \frac{\text{mechanical advantage}}{\text{velocity ratio}}.$$

Find the efficiency of the arrangement used in Expt. (iii) p. 306.

COMBINATIONS OF PULLEYS

If the number of pulleys is increased we can increase the mechanical advantage of the lifting-tackle. In the arrangement shown in Fig. 163 the load is supported by 6 strings, hence (neglecting friction and the weight of the movable pulleys) the pull

in any one string is one-sixth of the load; hence the force F required to hold the load in place is one-sixth of the load, and the theoretical mechanical advantage is 6. In order to raise the load through 1 ft. we have to pull in 6 ft. of rope at F, hence the velocity ratio is 6. (N.B.—The velocity ratio and the *theoretical* mechanical advantage of any such system of pulleys can be found by counting the number of strings passing through the *lower* pulley-block.)

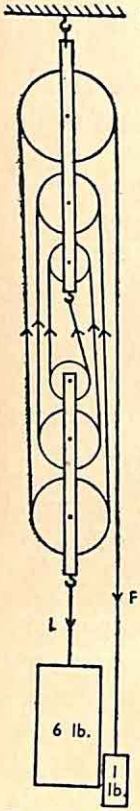


FIG. 163. Lifting tackle.

Set up the arrangement shown in Fig. 163.

(iv) First balance the hanging pulley-block with weights at F. Then hang a weight of 6 lb. from the lower pulley-block and find what force is, in addition, required at F to balance this load. $F =$ .....lb. wt. Therefore, mechanical advantage = ...

Verify the velocity ratio by measurement as in Expt. (ii) above = ... When the force F descends through 6 ft. *the work put into the machine* = force × distance moved by effort = ... foot-pounds, and *the work got out of the machine* = load × distance moved by load = ... foot-pounds.

(v) Now remove the weights used to counter-balance the hanging pulley-block and find what weight is required at F to balance the system. $F =$ ..... lb. wt. The *total work put into the machine* when the force F descends through 6 ft. = ...×6 =..... foot-pounds. The '*useful*' *work got out of the machine* during the same time = 6×1 = 6 foot-pounds.

$$\text{Efficiency} = \frac{\text{useful output}}{\text{total input}} = \frac{6}{\quad} \times 100 = \ldots\ldots \text{ per cent.}$$

THE SIMPLE WHEEL AND AXLE

This consists of a wheel of large diameter attached to an axle of smaller diameter. A similar machine, with a long handle in place

of the wheel (as shown in Fig. 164) is sometimes used for raising the water-bucket from a well.

The principle of the wheel and axle is shown in Fig. 165. Ropes

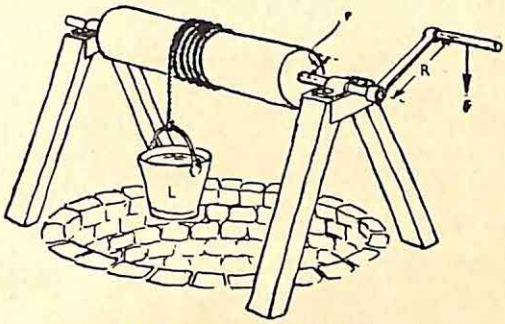


FIG. 164. Wheel and axle.

are wound in opposite directions round the wheel and the axle, the rope round the axle B carrying the load L, while the force F is applied to the rope wound round the wheel A. In one complete revolution of the wheel and axle, *the work PUT INTO the machine = force × circumference of wheel*, and *the work GOT OUT of the machine = load × circumference of axle.*

Therefore, *by the Principle of Work,*

$$\textit{load } (L) \times \textit{circumference of axle} = \textit{force } (F) \times \textit{circumference of wheel.}$$

Therefore,

$$\text{Mechanical advantage} = \frac{\text{load}}{\text{force}} = \frac{L}{F} = \frac{\text{circumference of wheel}}{\text{circumference of axle}} = \frac{\text{radius of wheel}}{\text{radius of axle}} = \frac{R}{r}.$$

During one complete revolution of the wheel and axle, the force F descends a distance equal to the circumference of the wheel and the load L rises through a distance equal to the circumference of the axle.

Therefore,

$$\text{Velocity ratio} = \frac{\text{circumference of wheel}}{\text{circumference of axle}} = \frac{\text{radius of wheel}}{\text{radius of axle}} = \frac{R}{r}.$$

Set up the arrangement shown in Fig. 165.

(*a*) Hang a weight of 4 lb. from the string wound round the axle

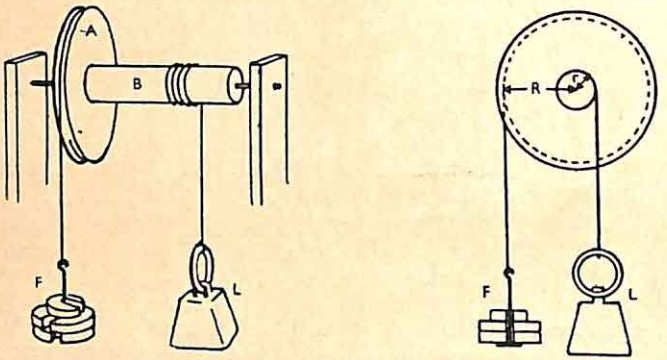


FIG. 165. Principle of wheel and axle.

and hang weights from the string wound round the wheel until it descends slowly but steadily. Force $F =$ ..... lb. wt.

(*b*) Pull down the string wound round the wheel until the load L rises 4 in. exactly and measure the distance through which the force F has descended. Load rises ..... in. Force descends ..... in.

$$\text{Efficiency} = \frac{\text{useful output}}{\text{total input}} = \frac{\text{load} \times \text{distance load moves}}{\text{force} \times \text{distance force moves}}$$

$$= \frac{\times}{\times} 100 = \text{.....} \text{ per cent.}$$

THE DIFFERENTIAL WHEEL AND AXLE

Since the mechanical advantage of a simple wheel and axle $= \frac{\text{radius of wheel}}{\text{radius of axle}}$, by having a very large wheel (or a very long handle) and a very thin axle, a very small force would lift an

enormous load *if the thin axle would stand it without breaking*. The same result is obtained with the *differential*† *wheel and axle* shown in Fig. 166, where the axle has two halves of different thicknesses

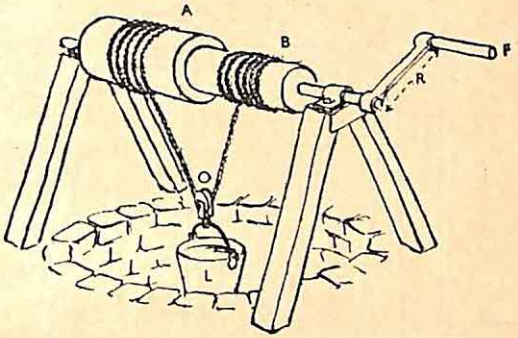


FIG. 166. Differential wheel and axle.

and the rope *winds up* round the thicker part A (of radius r_1) as it *unwinds* from the thinner part B (of radius r_2). Hence, in one complete revolution of the machine, the force F moves through $2\pi R$ (where R is the radius of the handle) and the rope shortens by $2\pi r_1 - 2\pi r_2$, i.e. the difference between the circumferences of A and B. But as the load L hangs from a movable pulley, it rises only *half* the length of rope wound up.

Hence, *work done on load* $= L \times \dfrac{2\pi(r_1 - r_2)}{2}$

$= L\pi(r_1 - r_2)$.

And *work done by force* $= F \times 2\pi R$.

Hence, *by the Principle of Work*, $L \times \pi(r_1 - r_2) = F \times 2\pi R$.

Therefore,

velocity ratio (and *theoretical* mechanical advantage L/F)

$$= \frac{2\pi R}{\pi(r_1 - r_2)} = \frac{2R}{r_1 - r_2}.$$

Set up a differential wheel and axle and find its actual mechanical advantage and its efficiency.

The same principle is used in the *differential pulley-block* shown in Fig. 167. This consists of two pulleys of slightly different radius,

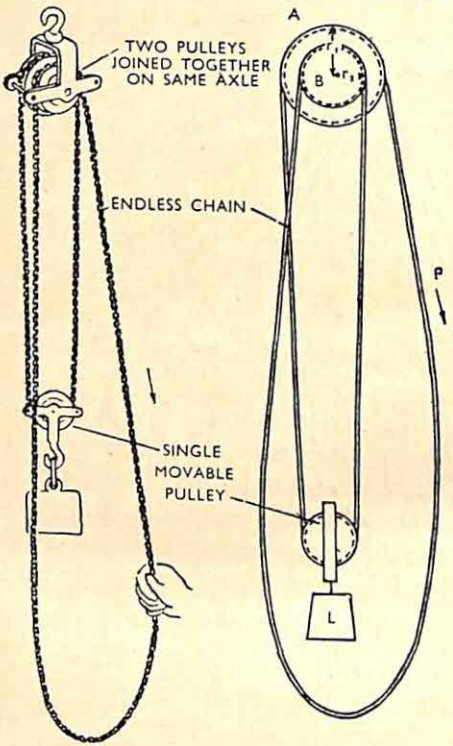


FIG. 167. Differential pulley block.

r_1 and r_2, *joined together* on the same axle, with an endless chain passing first over one fixed pulley, then round the lower movable pulley (which carries the load) and then up again over the other fixed pulley. Both the fixed pulleys have projections to stop the chain from slipping. As in the differential wheel and axle, during one complete revolution of the fixed pulleys the chain between the upper and lower pulleys is shortened by the difference between the circumferences of the two fixed pulleys, i.e. by $2\pi r_1 - 2\pi r_2$, and

the load is raised by *half* this amount. Hence, *the work done on the load L* during one complete revolution of the fixed pulleys is

$$L \times \frac{2\pi r_1 - 2\pi r_2}{2} = L \times \pi(r_1 - r_2),$$

and *the work done by the force* $= F \times 2\pi r_1$ (since the force F moves a distance equal to the circumference of the larger fixed pulley). Hence, *by the Principle of Work*, $L \times \pi(r_1 - r_2) = F \times 2\pi r_1$.

Therefore,

velocity ratio (and *theoretical* mechanical advantage L/F)

$$= \frac{2\pi r_1}{\pi(r_1 - r_2)} = \frac{2r_1}{r_1 - r_2}.$$

Hence, the smaller the difference in diameter between the two fixed pulleys, the greater is the mechanical advantage of the tackle.

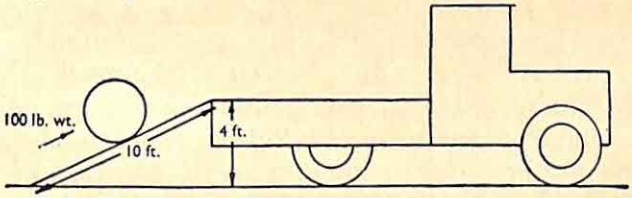


FIG. 168. Inclined plane.

If a differential pulley-block is available, find its actual mechanical advantage and its efficiency.

The efficiency of this type of machine is always less than 50 per cent., but this low efficiency is actually a practical advantage, because the load does not fall when the force is removed, as in ordinary lifting tackle (whose efficiency is over 50 per cent.).

THE INCLINED PLANE

Suppose that a man wants to put a heavy drum of oil, weighing 100 lb., into a lorry 4 ft. from the ground. It will be necessary to lift a weight of 100 lb. to a height of 4 ft., and this will require 400 foot-pounds of work. An easier way of putting the drum on the lorry will be to roll it up sloping pieces of wood placed against the back of the lorry to form an *inclined plane*, as shown in Fig. 168.

Suppose the length of this inclined plane is 10 ft. By the *Principle of Work*, it is necessary to do 400 foot-pounds of work, no matter what method is used, hence by using an inclined plane 10 ft. long,

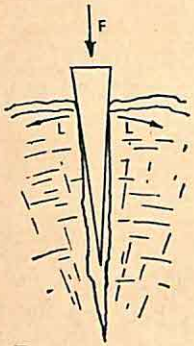


FIG. 169. Wedge.

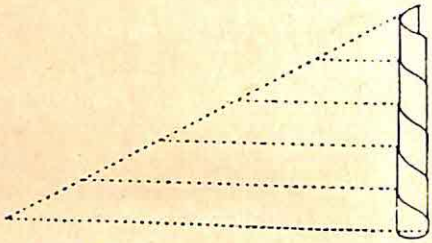


FIG. 170. A spiral 'inclined plane'.

the man can get the drum into the lorry by exerting a force of 40 lb. wt. over a distance of 10 ft., which will be much easier than exerting a force of 100 lb. wt. over a distance of 4 ft. by a direct lift.

This method of getting *mechanical advantage* is also used in *wedges* for splitting wood or stone, in *axes* and in *knives*. Suppose the wedge shown in Fig. 169 is 10 in. long and 2 in. wide at its flat end. For every 10 in. it is driven forward into a piece of wood, the wood is forced outwards 2 in. By the *Principle of Work*, this outward thrust L will be five times as great as the forward thrust F (neglecting friction). The efficiency of such a wedge, however, is usually less than 50 per cent. (fortunately for the user).

THE SCREW AS A MACHINE

Screws have many practical applications for getting a mechanical advantage. A screw combines the principle of the inclined plane with that of the lever. To show that a screw is a spiral inclined plane, cut a rectangular sheet of paper along a diagonal and wrap one of your right-angled triangles round a pencil as shown in Fig. 170. You will notice that the hypotenuse of the triangle forms a spiral '*screw-thread*' around the pencil.

In any screw the distance from one thread to the next, measured along the length of the screw, is called the *pitch* of the screw; this is the distance that the head of the screw moves forward into a fixed *nut* when the screw-head is turned through one complete revolution.

In Fig. 171 the force F acts at the end of the handle B (of length r) and raises a load L. During one complete turn of the screw through the fixed block A, the work done by the force $= F \times 2\pi r$, and the work done on the load $= L \times p$ (where p is the pitch of the screw).

Hence, by the *Principle of Work*, $F2\pi r = Lp$.

Therefore,

velocity ratio (and *theoretical* mechanical advantage L/F) $= \dfrac{2\pi r}{p}$.

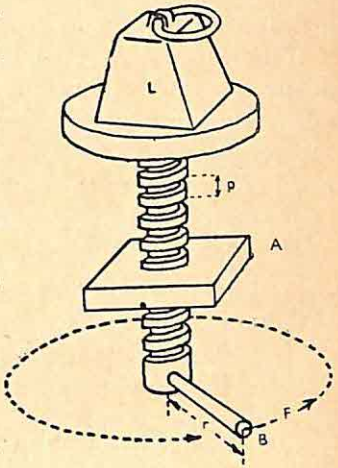


FIG. 171. Principle of screw-jack.

Hence, if the pitch is small and the handle is long an enormous mechanical advantage can be obtained. For example, a load of 2,112 lb. wt. rests on the head of a *screw-jack*† having 4 threads per inch (i.e. of $\frac{1}{4}$ in. pitch) turned by a handle 12 in. long. During one complete turn of the screw the load of 2,112 lb. wt. is raised through a height of $\frac{1}{4}$ in. ($= \frac{1}{48}$ ft.), so that 44 foot-pounds of work are needed (if there is no friction). By the *Principle of Work*, this is equal to the work done on the handle of the screw-jack, i.e. to a force of F lb. wt. acting through a distance of $2\pi \times 12$ in. $= 6{\cdot}28$ ft. Hence, $6{\cdot}28\ F = 44$, i.e. $F = 7$. Hence, with a 'perfect', frictionless screw, a force of 7 lb. wt. acting at the end of the 12 in. handle could raise a weight of 2,112 lb.

In actual practice, however, friction is very great in screws and a much greater force than 7 lb. wt. would be necessary to raise the load. In fact a frictionless screw would be of no use in a screw-jack because when the effort was removed the load would make the

screw untwist and so lose the height it had already gained. In practice we use the friction between screw and nut to hold the load when the force is removed. Consequently, the efficiency of a screw-jack may be as low as $\frac{1}{4}$, owing to the work wasted by friction.

If a motor-car screw-jack is available, find its actual mechanical advantage and its efficiency. A suitable 'load' will be a friend of known weight standing or sitting on a board fitted to the head of the screw-jack. Use a spring-balance to measure the force that must be applied to the ends of the handle to raise the load. Apply this effort at right angles to the handle and take the measurement *while the handle is moving* (you will find that the effort required to *start* the handle moving will be greater than this). Then remove the 'load' and find how far the head of the jack moves forward during 10 complete turns of the handle.

What is the velocity ratio of the jack?.....

What is its actual mechanical advantage?.....

What is its efficiency?.....

Many practical screw-jacks make use of the *differential* principle by having a *compound screw*, one part of which has a *right-hand thread* and screws *upward* while the other part has a *left-hand thread* and screws *downward*. The right-hand screw is of slightly larger pitch than the left-hand screw, hence the arrangement is mechanically equivalent to a single right-hand screw *of very small pitch* (which would have a very fine, weak thread) but the threads of the two parts of the compound screw can be made sufficiently strong to bear a heavy load.

(See *General Science Workbook*, Book Two, pp. 87–108, for numerical problems and learning exercises on this chapter.)

Chapter XIII

TIME AND SPACE

The measurement of time is of great practical importance in every-day life and in scientific work. In fact, it is possible to state all scientific measurements in the three fundamental units of *length*, *weight*, and *time*. Physical measurements are usually based on the *C.G.S. system* (*centimetre-gram-second system*), and engineering measurements on the *F.P.S. system* (*foot-pound-second system*).

The Earth rotates very steadily on its axis and it is therefore used as our standard timekeeper, from which we check all our clocks and watches. An observer takes the time at the moment when the Sun is due north of him, and again (one *solar day*† later) when it is due north after one rotation. The average length of all the *solar days* in a year is a *mean solar day* (or 'clock day') and this is divided into 24 *hours*. Each hour is divided into 60 *minutes* and each minute into 60 *seconds*.

The Earth also rotates round the Sun in an almost circular path (or *orbit*†). It takes 365·2425 *mean solar days* for the Earth to complete one revolution, and this is the *average value* of our *civil years*. Three successive civil years consist of 365 days and the fourth (or *leap year*) has 366 days; but century years do not count as a leap year unless divisible by 400.

Only *astronomers* can tell the time accurately by this 'earth-clock', and they control the 'master-clocks' from which we get our *standard time*. For example, the wireless *time-signal* sent out at the same time every day from Greenwich Observatory* marks a mean solar day of 86,400 seconds and its accuracy never varies by more than 0·01 second per day. Nowadays, accurate time can be kept by clocks and watches adjusted to these daily wireless time-signals, but centuries before clocks and watches were invented there were other less accurate methods of measuring time. For

example, people used *water-clocks*, *hour-candles*, and *hour-glasses*. In the simplest type of water-clock, water was allowed to drip slowly (but regularly) from one vessel to another. The rise of water in the lower vessel could then be used to measure time, e.g. if the level rose 24 in. in one day, then each rise of 1 in. represented 1 hour. Hour-candles burnt at a steady rate and were marked off in units, each unit being the length of candle that burnt away in an hour. In sand-glasses (or *hour-glasses*) fine sand ran through a narrow passage between two glass bulbs, taking 1 hour to empty the upper bulb (see Fig. 172).

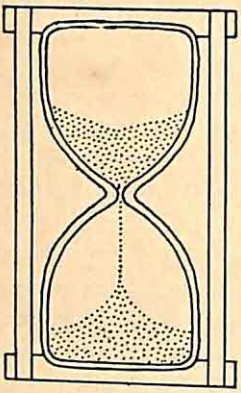


FIG. 172. Sand-glass.

SUN-DIALS

One of the earliest instruments for measuring time was the *sun-dial*.† If a stick is fixed upright in the ground, when the Sun rises in the east the stick casts a long shadow westwards, and this shadow moves round the stick until, when the Sun sets, the shadow points eastwards. If a mark is made on the ground in the path of this moving shadow, the shadow will fall on the mark once each day and the time-interval between two successive shadows passing this mark is a *solar day*. The length of this time-interval is slightly different from day to day, varying with the distance of the Earth from the Sun as it follows its path round the Sun along an *ellipse** and not in a circle. Only four times each year—on 15 April, 15 June, 1 September, and 24 December —are the *solar day* and the *clock-day* of the same length. The greatest difference between them is about 16 minutes at the beginning of November.

In a practical sun-dial, like the one shown in Fig. 173, which is specially suitable for places near the Equator, the rod that casts the shadow forms the axis of a cylinder, the bottom half of which catches the shadow. The *shadow-rod* points due north and south

and it is inclined to the horizontal at an angle equal to the *latitude** of the place where the sun-dial is to be used. That is, the axis of the cylinder (the shadow-rod) is *parallel to the Earth's axis*. Noon

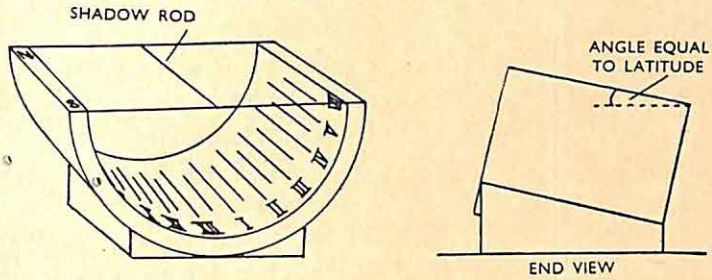


FIG. 173. Cylindrical sun-dial for use near Equator.

is marked vertically below the shadow-rod, the horizontal edges of the half-cylinder are marked 6 a.m. and 6 p.m., and 12 equal intervals (hours) are marked off round the half-cylinder.

Such a sun-dial gives *sun-time* (or *local apparent time*) and this agrees with *clock-time* only on four days each year. Hence a table of corrections is needed to get clock-time from the sun-time shown by the sun-dial.[1] (In countries with 'daylight-saving' laws, a further correction is necessary.) A sun-dial, however, is useless at night or in cloudy weather.

None of these early methods for measuring time was of any use for accurate measurement of time for scientific purposes, and accurate time-measurement had to wait until the discovery of the properties of the *pendulum* in the sixteenth century.

Galileo, who laid the foundations of physics with his experiments on moving bodies, was also responsible for the earliest observations of the properties of *pendulums*. In 1583, it is said, he watched the swinging of a number of lamps hanging from the ceiling of a building. All the lamps hung by chains of equal length, and as

[1] TEACHER'S NOTE: This can be worked out from the 'Nautical Almanack'. (A concrete half-cylinder, 3 ft. in diameter, as used for large concrete drains, makes an excellent permanent sun-dial. If the table of corrections is worked out for ten-day intervals throughout the year, it will be sufficiently accurate for school purposes.)

they swung to and fro, Galileo noticed that the time of swing was the same for every lamp, although some were swinging through a much wider arc than others. (Galileo used his own *pulse* for these time-measurements.)

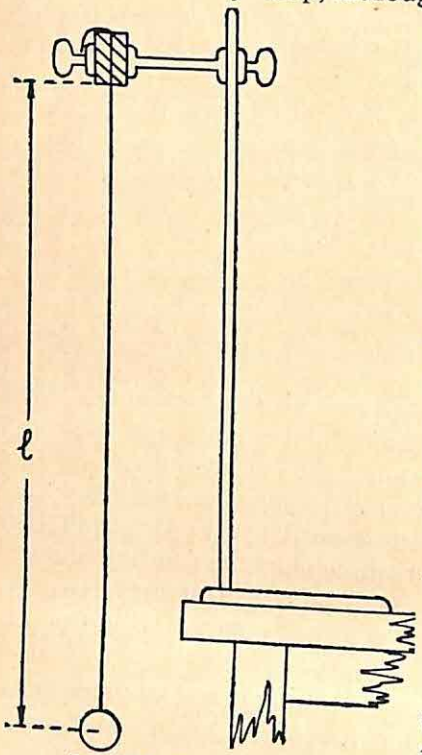


FIG. 174. Simple pendulum.

THE SIMPLE PENDULUM

An 'ideal' simple pendulum is defined as *a heavy particle suspended by a weightless thread.* The nearest we can get to this in practice is to hang a small metal ball or '*bob*' on a very thin thread, and we shall now use this arrangement to carry out experiments to illustrate the most important properties of the simple pendulum.

TO SHOW THAT THE TIME OF SWING IS INDEPENDENT OF THE ANGLE OF SWING

Thread a piece of cotton thread, about 2 metres long, through a small hole in a cork, and fix the cork in a retort-clamp attached to a stand (see Fig. 174). Hang a metal 'bob' to the end of the thread and adjust the length so that the *centre* of the bob is exactly 1 metre below the *bottom* of the cork. This distance is called *the length of the pendulum.*

Set the pendulum swinging by pulling the bob about 1 in. to the right. Then, with a stop-watch, take the time of 50 *complete swings*[1] (i.e. 'to-and-fro' or 'there-and-back'), being careful to

[1] Stand in front of the retort-stand, and count complete swings every time the bob passes the rod *when moving from right to left.*

count 'nought' as you begin timing and 'one' at the *end* of the first complete swing.

Time of 50 swings = ... seconds.

Repeat this, pulling the bob about 2 in. to the right, and again take the time for 50 wider swings.

Time of 50 swings = ... seconds.

Repeat this once more, pulling the bob about 3 in. to the right when you start the pendulum swinging, and again take the time of 50 still wider swings.

Time of 50 swings = ... seconds.

If your experiment is done carefully, you will find that *the time of swing is independent of the angle of swing* (*provided this is small*).

TO SHOW THAT THE TIME OF SWING IS INDEPENDENT OF THE WEIGHT OR MATERIAL OF THE BOB

Fit up two simple pendulums of exactly the same length (about 1 metre) but with bobs of different material and different weight, e.g. use lead, brass, steel, wood, or glass bobs. Pull both bobs about 3 in. to the right and let them go at the same moment. Notice that their times of swing are exactly the same. Hence, *the time of swing is independent of the weight or material of the bob.*

TO SHOW THAT THE TIME OF SWING DEPENDS ON THE LENGTH OF THE PENDULUM

Measure the time of 50 swings of a pendulum 1 metre long. Shorten the pendulum to 81 cm. and again find the time of 50 swings. Repeat this with pendulums 64 cm., 49 cm., 36 cm., and 25 cm. long, recording your figures on p. 322. Plot graphs (see p. 322) of (i) *length* (l) *against time* (t), (ii) $\sqrt{l}$ *against* t.

If your experiments have been carried out carefully all the figures in the last column will be the same, and your graph of $\sqrt{l}$ against t will be a straight line, showing that, for a simple pendulum, *the time of swing is proportional to the square root of the length of the pendulum.*

Length of Pendulum (l)	Time of one complete swing (t)	$\sqrt{}$Length ($\sqrt{l}$)	$\sqrt{l}/t$
cm.	$\div$ 50 = sec.	cm.	

PENDULUMS AND BALANCE-WHEELS

We see, therefore, that the time of swing of a pendulum is always the same provided that its length is not altered and that the angle of swing is small. In other words:

(*a*) *The time of swing is independent of the angle of swing* (*provided that this is small*).

(*b*) *The time of swing is independent of the weight or material of the pendulum-bob.*

(*c*) *The time of swing depends only on the length of the pendulum.*

Under suitable conditions a spring can be made to behave like a pendulum, always taking the same time for a complete to-and-fro movement.

Most modern clocks and watches are controlled by *pendulums* or by *balance-wheels* making use of this property of constant time of swing.

CLOCKS AND WATCHES

In mechanical clocks and watches the 'works' are turned either by a hanging weight or by a compressed spring. Left to itself, the weight or spring would make the hands of the clock spin round at a high speed, but every clock or watch has a device, called the *escapement*,† for controlling the speed of the wheels.

A simple type of escapement used in pendulum-clocks is shown in Fig. 175. This allows the fastest-moving wheel of the 'clockwork' to move forward from one tooth to the next as the pendulum completes each to-and-fro swing. At the same time, the escaping tooth of the escapement-wheel gives the escapement a slight push that keeps the pendulum swinging. In a tall 'grandfather' clock the pendulum is 39 in. long, and makes one half-swing every second. If its escapement-wheel has thirty teeth, this wheel (which carries the 'seconds-hand') will make one revolution per minute.

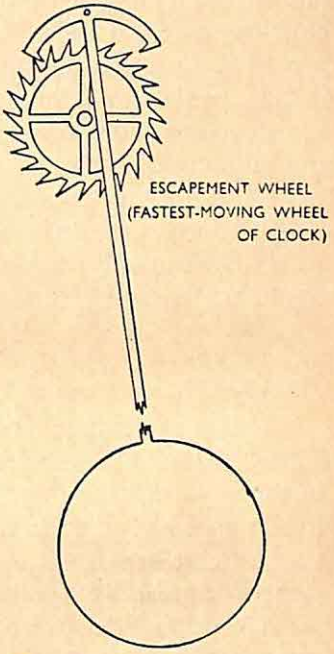


FIG. 175. Clock escapement.

In smaller clocks and in watches the *balance-wheel* serves the same purpose as the pendulum, carrying the escapement that controls the movement of the fastest-moving wheel in the 'clockwork'. The balance-wheel swings backwards and forwards under the action of a spiral *hair-spring* in much the same way as a pendulum swings under the pull of gravity.

A *balance-wheel*, in fact, is a *compound pendulum*, with the same property of always taking exactly the same time for a complete swing provided that the size and weight of the balance-wheel and the strength of the hair-spring do not alter. In pendulum-clocks the time of swing is adjusted by altering the length of the pendulum. In small clocks and watches the time of swing is adjusted by altering the length of the hair-spring, shortening it to make the wheel swing faster, and vice versa.

(See *General Science Workbook*, Book Two, pp. 94–98, for numerical problems on pendulums.)

THE EARTH IN SPACE

Astronomy—the scientific study of the heavenly bodies—is one of the most ancient of the sciences. For thousands of years Man has been making observations of the sky at night and drawing conclusions from them. Some of his knowledge he has put to great practical use, e.g. in navigation* and timekeeping; much of it has served only, as yet, to increase his desire to know more about the structure and working of the Universe. In recent years, with the building of better telescopes and with the increase in mathematical ability, astronomy has made great progress. We cannot, in this course, explain how the astronomer goes about his scientific work—we can only describe some of his more important discoveries about the Earth's position in space.

THE HEAVENS

The first astronomers began by naming and classifying the stars. They realized that most of them occupied unchanging positions in the sky *relative to one another*, forming patterns, or *constellations*,† to which they gave names like Scorpio and Orion. These star patterns have changed so very little over the last three thousand years that we still keep these names although many new constellations have been discovered with the aid of telescopes. With a little patience and the help of a star map you will find it easy to recognize the chief constellations.

When you know a few constellations, map their position at different times on the same night. Although they keep their *relative* positions, the star-field *as a whole* seems to move. Those constellations that were low in the east at nightfall rise higher in the sky as time passes, and others appear to rise from below the horizon. At dawn, the brightest of the stars—the Sun—appears over the eastern horizon. It is as though the stars were holes in a huge inverted bowl that is spinning round the earth with its axis of revolution passing through the *celestial poles*.† This is, in fact,

what ancient people believed. Later Man came to realize that it is the *Earth* that is moving, spinning on its axis from west to east. Extension of the Earth's (imaginary) axis indicates two positions in the sky about which the stars seem to revolve. In the northern hemisphere one of these poles is, as it happens, marked by the small N-pole star. In the southern hemisphere there is no such pole star, the actual celestial S-pole being near the constellation called the *Southern Cross*.

The clearest demonstration of this apparent movement of the heavens, due to the Earth's rotation, is given by a photographic time-exposure of about 8 hours, at night, taken with the camera pointing at one celestial pole. Such a photograph shows a series of concentric* star tracks—arcs of circles—centred at the pole. These tracks are made because the Earth carries the camera round in the arc of a circle against the background of fixed stars.

Your next observation of the night sky should be of the *planets* ('wanderers'). These are bright, star-like bodies that change their positions nightly against the general background of 'fixed' stars. Their apparent movements are not simple, e.g. for several months a planet may wander eastward among the constellations, then westward, then eastward once more. The planets also differ from the stars in shining with a steady light—they do not 'twinkle'.*

A much more obvious heavenly body is, of course, the *Moon*. If you observe it each evening at the same hour, you will notice that, apart from its apparent changes in shape, from 'new' to 'old', it, too, wanders against the starry background.

We now know that the movements of the planets and of the Moon are partly due to the Earth's rotation (i.e. they are only apparent movements like those of the stars) and partly due to the actual movement in the heavens of the planets and of the Moon. We know that the Earth is a great sphere, spinning on its own axis and at the same time revolving round the Sun, the star nearest to it. In their movements the planets resemble the Earth; they are a part of the *Solar System*.† The Moon is the Earth's *satellite*† (i.e. its orbit of revolution is round the Earth) and as it travels around

the Earth in its own fixed orbit it is carried along in the Earth's orbit round the Sun.

THE STARS

The stars are suns of various sizes, many millions of miles from us. Our Sun is actually rather a small star, about 100 times the Earth's diameter and 93 million miles away. Like the other stars, it is constantly giving out light and heat. During *eclipses* we can see that its surface is a great cloud of fiery gases whose temperature has been estimated at 6,000° C. (see Fig. 176).

The nearest star to us, next to the Sun, is 25 billion* miles away. The mile is too small a unit for measuring such enormous distances, and it is more convenient to use a special astronomical unit of length—the *light-year*.† Light travels at 186,000 miles per second. In a year it would travel 5,880,000,000,000 miles. This distance is a light-year. Our nearest well-known star, Sirius (the 'dog-star'), is 8·6 light-years away; that is, light given out from Sirius and travelling at 186,000 miles per second would take 8·6 years to reach the Earth!

There are millions of stars in the heavens, some larger, some smaller, some hotter, and some colder than the Sun. The biggest of them all is *Antares*,[1] 400 million miles in diameter. As seen from the Earth, most of the stars seem to form a circular belt in the heavens—the Milky Way. Astronomers now believe that a very large number of stars (perhaps 3,000 million) are grouped in a *galaxy*† (one of many known to exist) which is roughly lens-shaped and of enormous size (about 2,500,000 light-years in diameter). The solar system is nearer the centre than the edge of this galaxy, so when we look outwards from the Earth, most of the stars seem to be grouped in the girdle of the Milky Way.

One at least of the millions of stars in our own galaxy, the Sun, is surrounded by a number of planets. One of these planets is the Earth—a comparatively tiny particle in the vast spaces of the Universe.

[1] Antares is the reddish star at the 'heart' of the constellation Scorpio.

The Sun, with the bodies that revolve around it, is called the *solar system*; it consists of the Sun, the nine *planets* with their moons, the *planetoids*,† and some *comets*† and *meteors*.

The Sun is a very large, very hot ball sending out light and heat

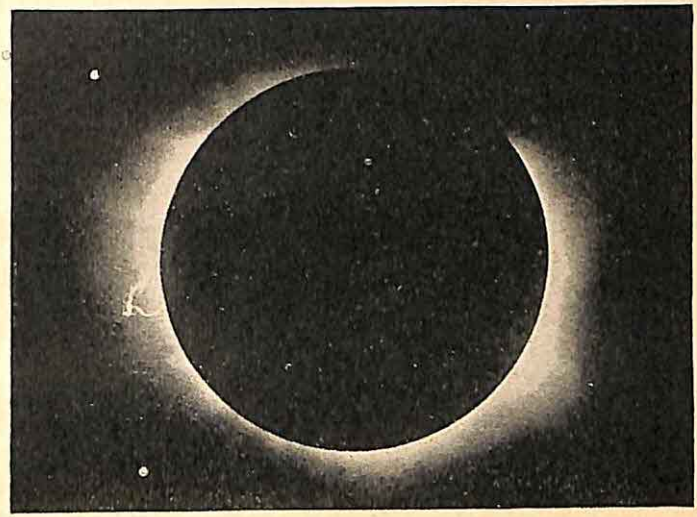


FIG. 176. Solar eclipse.

to the other members of the solar system. At the surface its temperature is about 6,000° C.; at the centre it is estimated to be about 40 million degrees. Yet as far as we can tell, the Sun is not being consumed* by any sort of combustion. It appears likely that in the Sun *atoms* are being continually broken up and re-formed, and it is from this atomic source that the Sun gets its energy. At normal times the Sun is so bright that we can only study its *light*, although small dark *sun-spots* become visible on the disc at regular intervals. However, during the occasional *total eclipses* of the Sun by the Moon, the light from the centre part of the Sun is cut off and the full beauty of the star can be realized. The eclipsed disc is then

seen to be surrounded by a bright red ring (containing the elements hydrogen, helium, and calcium). Rising from this are *prominences*,* several hundred thousand miles high, often brightly coloured.

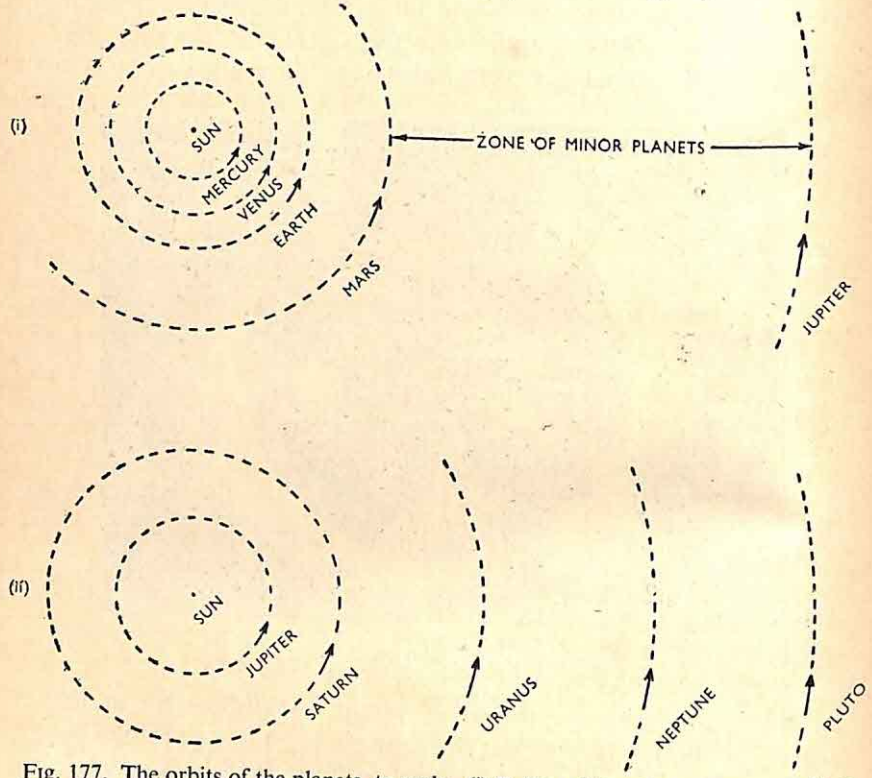


Fig. 177. The orbits of the planets, to scale. (i) Orbits of the planets nearest to Sun. (ii) Orbits of more distant planets. In order to show the more distant planets on the page, (ii) has been drawn about eight times smaller than (i).

These stretch out into the *corona*,† a ring of light formed by reflections from minute particles (see Fig. 176).

THE PLANETS

Although they appear to us about the same size as, or even larger than, the bigger stars, the nine planets are comparatively small.

They appear large to us only because they are relatively near. Also, they differ from the stars in that their light is merely *reflected* from the Sun. They have a solid crust and, although some of them are much hotter than the Earth, they radiate no light or heat of their own. They all rotate around the Sun in the same direction, from west to east, but at different speeds and in different orbits. Their orbits are nearly circular ellipses. Most of the planets are known to revolve on their own axis, from west to east. They contain the same chemical elements as the Sun, or, at least, some of them.

The planets nearest the Sun—*Mercury*, *Venus*, and the Earth—are comparatively small, and so is the recently discovered *Pluto*, the outermost of the nine. The biggest planet is *Jupiter*. In Fig. 177 are shown the orbits drawn to scale and some further details are given in Table I below.

Some idea of the relative size of the bodies in the solar system is given by the following illustration: let a globe 2 ft. in diameter represent the Sun. Mercury will be then represented by a grain of mustard seed on the circumference of a circle 164 ft. in diameter; the Earth by a pea on a circle of 430 ft. diameter; Jupiter by an orange on a circle of about $\frac{1}{2}$ mile diameter.

TABLE I

Planet	Equatorial Diameter (miles)	Distance from Sun (millions of miles)	Period of Revolution	Number of Moons
Mercury	3,000	36	88 days	0
Venus	7,580	67	225 ,,	0
Earth	7,926	93	365¼ ,,	1
Mars	4,220	141	1·88 years	2
Jupiter	88,700	483	11·86 ,,	11
Pluto	2,000	3,680	249·2 ,,	—

The other planets are *Saturn*, *Uranus*, and *Neptune*.

Venus is a planet not unlike the Earth in size and mass. It is near the Sun and therefore visible in the sky just after sunset and

before sunrise; at favourable times it is easily the brightest object in the sky and may even be seen in broad daylight. It shows *phases*,* like the Moon. It is our nearest planetary neighbour, but we know little about it, because its surface is always hidden by a cloudy atmosphere. There *may* be living things on Venus, but we have no evidence of their actual existence.

Mars has a diameter about half that of the Earth. At its nearest approach to us it is 35 million miles away. Yet with modern powerful telescopes and other apparatus astronomers are able to learn a good deal about its surface. Its atmosphere probably contains water-vapour and oxygen, and although the atmosphere of Mars is much less dense than ours, it is *possible* that living things exist on Mars. The surface of the planet shows areas of different colours, and some of these areas change in size with the seasons, which at least *suggests* that they may be covered with plants. It is of interest that the polar regions of Mars, like those of the Earth, are covered with snow or ice (or with something that looks very like it), and that the area of the polar caps also increases and decreases with the seasons.

THE MOON

The *Moon* is a satellite of the Earth around which it revolves once every 27 days, in a circle of only 240,000 miles radius. The Moon's diameter is only one-quarter that of the Earth. Its light is reflected from the Sun. Only one-half of the Moon's surface is illuminated at once and it presents to the Earth the dark side (new moon), the light side (full moon), or part of the light side (before and after full moon).

The Moon has no atmosphere and certainly carries no life. It is a dead world of remarkable peaks,* craters,* and desert-plains. It is also an unchanging world, for there is no erosion taking place on its surface.

The chief cause of the *tides* on the Earth is the gravitational attraction of the Moon on the mass of water of the oceans. The waters become heaped up immediately under the Moon and also

(for reasons that we cannot give here) on the side of the Earth opposite to this. These two 'heaps' of water remain still, relatively to the Moon, while the Earth rotates. Hence two high tides and two low tides occur each day at any one place. At full moon and new moon the Sun and the Moon and the Earth are in a straight line and the tidal action is therefore increased owing to the Sun's gravitational attraction. Hence at these times we get very high and very low *spring tides.*

PLANETOIDS, COMETS, AND METEORS

The planets are not the only bodies that revolve round the Sun. Between the orbits of Mars and Jupiter are over a thousand tiny planets, the *asteroids*† or *planetoids* that we cannot see with the naked eye; none of these approaches the Earth in size. More remarkable than these are the *comets*, a few of which are of great brightness. Most of them are gaseous bodies of small mass. Their orbits are elliptical and they only appear near the Earth at comparatively infrequent periods that may be calculated if the orbit is known. Thus the famous Halley's Comet has a period of 75½ years and last appeared in 1910. Large comets of this kind have an extremely bright nucleus, like a planet, surrounded by a luminous ring (or *halo*†) and with a great tail, sometimes stretching across half the sky. In the past the comets have been regarded with fear. Actually they are largely gaseous and they shine by light reflected from the Sun.

Meteors are much more familiar to us—as 'shooting stars'. They are mostly tiny particles of matter—of stone or metal (e.g. iron and nickel). It is possible that some are the remains of old comets, and these appear in 'showers' at regular intervals. Others are wandering visitors from interplanetary space. When they reach the Earth's atmosphere, at a height of about 100 miles and moving at about 30 miles per second, friction with the atmospheric gases is sufficient to make them white-hot. The great majority burn away before they reach the Earth's surface, but some do reach the Earth's crust. The largest known to do so was 50 tons in weight.

Meteors seldom cause damage, although in 1908 a shower of large ones caused great destruction in Siberia.

THE ORIGIN OF THE PLANETS

The planets all revolve in the same direction that the Sun turns on its axis and their orbits are all in one plane. It is believed that they were all originally part of the Sun, and that they were dragged out of the Sun by the gravitational attraction from a passing star, as a thread of exceedingly hot matter that later broke up and formed liquid, white-hot, planets with their moons. These planets gradually cooled and formed solid crusts. They continue to revolve around the Sun, whose *gravitational attraction exactly counterbalances their tendency to fly off in straight lines out into space.*

Section IV
ELECTRICITY

Chapter XIV
ELECTRIC CHARGES AND ELECTRIC CURRENTS

Until the end of the eighteenth century continuous electric *currents* were unknown, although electric *charges* and momentary electric *shocks* and *sparks* had been produced by rubbing different substances together. For example, when dry glass is rubbed with dry silk, one kind of electric charge appears on the glass and another kind of charge on the silk, and two bodies can be charged in opposite ways until a momentary electric spark passes between the two charged bodies when they are brought near together. This '*frictional electricity*' (or *static* electricity), however, is not a suitable source of power for everyday purposes, although it produces important effects in Nature, e.g. lightning. Electrical *current* is used in all our modern electrical devices, e.g. lamps, dynamos, and electric motors. This is produced in quite a different way, but, as you will see later, it has the same essential properties as frictional electricity.

STATIC ELECTRICITY (ELECTROSTATICS)

Electrically charged objects exert a force of attraction on other objects. Such objects, under suitable conditions, may be able to move and *do work*, showing that electricity is a form of energy.

ELECTRICITY PRODUCED BY FRICTION (RUBBING)[1]

(1) Rub an ebonite or sealing-wax rod with flannel* or catskin,

[1] Experiments on frictional electricity are difficult to carry out unless the atmosphere is dry.

then hold it near small pieces of paper, straw, hair, gold-leaf, etc. The rod is said to be 'excited': it is charged with *static* or *frictional electricity*.

(2) Repeat the experiment, using a clean warm glass rod rubbed with silk.

(3) Balance a wooden ruler on a paper stirrup hanging by a thread or on an inverted watch-glass, then bring a charged ebonite rod near one end of it.

Charges of electricity are produced by friction and these remain on ebonite, silk, sulphur, sealing-wax, etc. These charges attract and move nearby light bodies. Rods of copper, iron, or other metal, when held in the hand during rubbing, do not acquire this power of attraction. Later on we shall discover why.

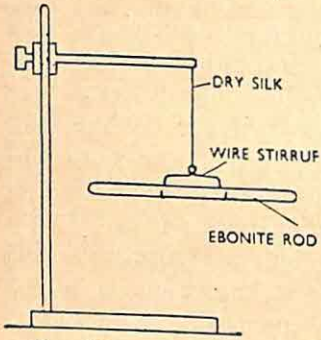


FIG. 178. Electric charges.

THE TWO KINDS OF ELECTRICITY—THEIR ATTRACTION AND REPULSION

(1) Rub an ebonite rod with flannel to charge it and then balance it in a copper wire stirrup as shown in Fig. 178. Bring near one end of it another charged ebonite rod. They repel each other. Repeat the experiment, using two charged glass rods. They also repel each other. So we conclude that *like* charges of electricity *repel* each other.

(2) Balance an electrified ebonite rod as before. Then bring near it a charged glass rod. They attract each other.

We can explain these observations if we assume that there are two kinds of static electricity, and that *unlike charges attract*, while *like charges repel*, one another. For convenience we call the one kind *positive* (as on glass) and the other *negative* (as on ebonite).

CONDUCTORS AND NON-CONDUCTORS OF ELECTRICITY

A *conductor* is a substance that allows electric charges to move along it: metals are *good conductors*. (When a rod of copper is rubbed, the charges produced by friction do not remain there but flow along the copper into the person holding the rod.) Water is a conductor, and so is damp air. (This will explain why it is difficult to carry out these experiments on static electricity in the wet tropics where the humidity is always high.)

A non-conductor (or *insulator*) is a substance that prevents the flow of electricity, e.g. ebonite, glass, sulphur, rubber, porcelain, dry air.

HOW TO DETECT ELECTRIC CHARGES

(*a*) *The pith ball electroscope*

An *electroscope*† is an instrument for detecting small charges of electricity and for determining the kind (or sign) of an electric charge. A simple instrument of this kind can be made by suspending two small light balls of dried pith on dry silk threads (non-conductors) from a support (see Fig. 179).

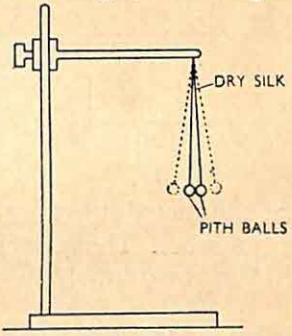


FIG. 179. Pith-ball electroscope.

Bring a negatively charged ebonite rod up to the pith balls. Notice that at first they move towards the rod, just as does a small piece of paper. As soon as they touch the rod, however, they move away from it and *fly apart*. Each ball has become negatively charged by contact with the rod and, as like charges repel, they diverge.

Now bring a negatively charged ebonite rod up underneath the two pith balls: they diverge still more. Why? Bring up next a positively charged glass rod. What happens? Why?

(*b*) *The gold-leaf electroscope*

A much more sensitive instrument is the *gold-leaf electroscope*,

in which the pith balls are replaced by two metal strips, one of which is an extremely thin and light leaf of gold. A simple electroscope of this type is shown in Fig. 180.

The electroscope is made by soldering one end of a brass rod A to a circular brass plate B. The other end of the rod is beaten out to form a flat plate C. The rod is supported as shown, with the plate C inside a glass bottle. The rod A passes through an insulating plug S of sulphur in the cork of the bottle. (This plug is prepared by just melting some sulphur and pouring it into the large hole in the cork while the rod is held in position.) The gold leaf G is fixed parallel to C by means of a tiny drop of gum on its upper edge.

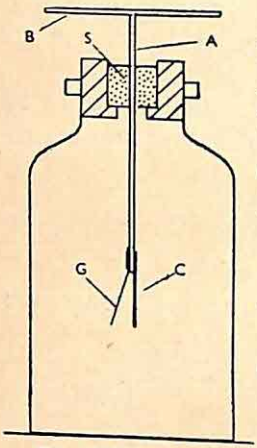


FIG. 180. Gold-leaf electroscope.

Bring a weakly charged ebonite rod in contact with the brass plate of the gold-leaf electroscope; then remove the rod. The leaf G diverges from C and remains apart. Negative charges have flowed from the ebonite into the gold-leaf electroscope and have been conducted down into G, which diverges from C because like charges repel. The gold-leaf electroscope is very sensitive because gold leaf is extremely thin and light.

ELECTROSTATIC INDUCTION

We now know that unlike charges of electricity attract each other, and a further experiment will show us why a charged rod will attract small light objects (e.g. a pith ball).

Cover a short wooden cylinder (say 3 in. long by $\frac{3}{4}$ in. diameter) with tin foil, so that its surface becomes a conductor. Place it on an insulating (non-conducting) stand, e.g. of paraffin wax or sealing wax. Now charge a single suspended pith ball negatively by touching it with a charged ebonite rod. Test each end of your tin foil-covered cylinder for charge by bringing up the pith ball;

it is weakly attracted by each end. Repeat the test with a *positively* charged pith ball; at each end of the cylinder this, too, is weakly attracted. The cylinder is therefore uncharged or neutral.

Now bring *near* to the end of the cylinder a charged ebonite

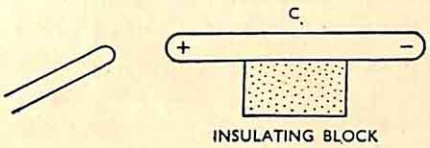


FIG. 181. Electrostatic induction.

rod. Keeping it in position, test each end of the cylinder with the pith balls. Show that the cylinder is now charged as indicated in Fig. 181.

The ebonite rod has *induced* a positive charge at the left-hand end of the cylinder and an (equal) negative charge at the opposite end. We believe that what happens is this: in the original neutral conductor C both charges exist, equally distributed. When the negatively charged rod is brought near to one end it repels the negative charges, which pass along the conductor, away from the rod, leaving that end positive and making the opposite end negative.

Now suppose the cylinder C to be very light. The negative charge on the ebonite and the positive charge on the left-hand end of C would attract one another and the cylinder would move towards the rod. Hence electrostatic induction explains the attraction between charged bodies and small, light, neutral (conducting) bodies.

The gold-leaf electroscope is usually *charged by induction.* The method is illustrated in Fig. 182 (*a*). To charge the electroscope positively, bring near to the disc B a negatively charged ebonite rod. This induces an opposite positive charge on B and a negative charge on G and C, so that G diverges from C.

While still keeping the ebonite rod in position, touch the plate B, when the negative charges will pass to earth and the leaf G will collapse, see Fig. 182 (*b*). Remove the finger; the positive charge

remains on the plate, being attracted to the opposite charge on the ebonite rod, see Fig. 182 (*c*). Now remove the rod; the positive charge becomes equally distributed over the electroscope and the leaf diverges again, see Fig. 182 (*d*). The electroscope has been positively charged by induction and is now a convenient instrument for testing the sign (+ve or −ve) of other charged bodies.

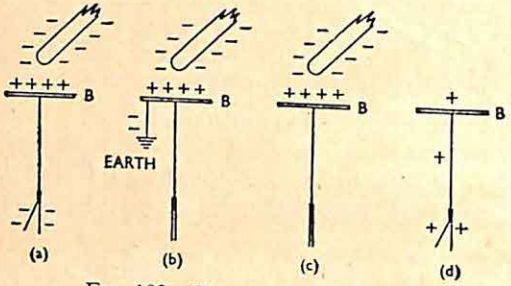


FIG. 182. Charging by induction.

(It may, alternatively, be charged negatively by means of a positively charged glass rod.)

LIGHTNING

Static electric charges are produced in large storm clouds, probably during the rapid upward air movements that tear apart the falling rain-drops and leave the droplets electrically charged. In time, very high charges are built up on the cloud. These can induce charges on another cloud or on a building and eventually a current of electricity may pass between the two, neutralizing the charges. This current, passing through air (which is a bad conductor), heats it and produces a bright flash of *lightning*. The heated air expands and sends out a pressure wave that we hear as *thunder*. If such electrical discharges take place through the human body they may cause death; if through trees or buildings (which are better conductors than animals) they often cause great damage because of the heating effects. To allow the charged cloud to be discharged easily and without danger to the building, *lightning arrestors* or conductors are fitted on tall buildings. The electricity

then travels to earth through a strip of metal (a good conductor) instead of through the building (a bad conductor).

CURRENT ELECTRICITY

If there is an unlimited supply of electric charges to a conductor, there will be a *continuous flow of electricity* in a definite direction along the conductor. This is called an *electric current*. Continuous

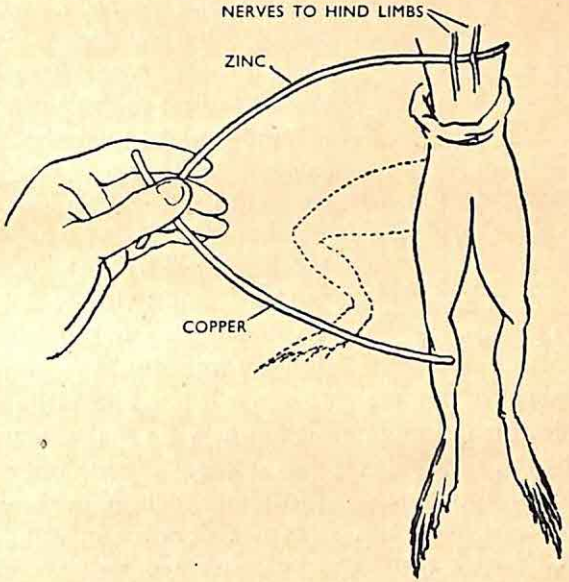


FIG. 183. Galvani's historic experiment.

electric currents were unknown until the end of the eighteenth century. The first step towards the production of a continuous electric current was made accidentally by Galvani, an Italian scientist, in 1786, while he was carrying out experiments on the nerves and muscles of frogs, using a small electric *shock* as a stimulus to produce muscular contraction. One day he got a similar result by holding a piece of copper and a piece of zinc in contact between his fingers, allowing the other end of the copper to touch a nerve and the other end of the zinc to touch a leg muscle, as shown in Fig. 183. Galvani explained this by saying that the

frog's muscles contained electricity, but Volta, another Italian scientist, thought that the electricity that made the frog's legs kick came from the contact between the two different metals, copper and zinc. Volta experimented with pairs of different metals and, in 1800, he showed that *when two different metals are separated by a solution that reacts chemically with one of them, an electric current flows along a wire joining the two metals.* He made the first *voltaic*† *cell* from discs of zinc and copper separated from each other by discs of porous paper soaked in salt solution.

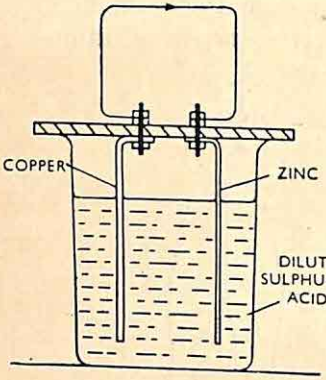


FIG. 184. Simple cell.

THE SIMPLE VOLTAIC CELL

The simplest voltaic cell consists of a sheet of *zinc* and a sheet of *copper* dipping into dilute sulphuric acid.

When a sheet of commercial zinc is dipped into dilute sulphuric acid, bubbles of hydrogen appear, showing that the zinc is being dissolved by the acid. This is called '*local action*' and is caused by the presence of impurities, e.g. iron and carbon, in the commercial zinc. *Chemically pure zinc* does not dissolve in dilute sulphuric acid. This 'local action' can be prevented by covering the zinc plate with a thin layer of mercury, which dissolves some of the zinc but not the impurities, forming a bright coating that consists of a solution of pure zinc in mercury. If sheets of mercury-coated zinc and copper are put into a beaker of dilute sulphuric acid, neither metal dissolves *until the copper touches the zinc*, when the zinc begins to dissolve. (This should remind you of your preparation of hydrogen, when you added some copper sulphate solution to deposit a little copper on the zinc and make it dissolve, see Book One, Chap. V.)

When the lower halves of the zinc and copper plates are separated from each other by dilute sulphuric acid, as shown in Fig. 184, and

their upper ends are joined by a wire, *a current flows through the wire*. The direction of the current *in the wire* is said to be *from the copper towards the zinc*; and it flows *from the zinc towards the copper through the acid inside the cell*. The *zinc* is called the *negative pole* of the cell and the *copper* is called the *positive pole*.

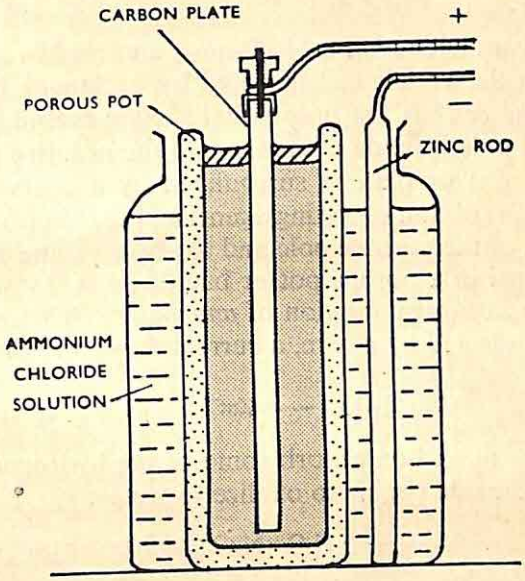


FIG. 185. Leclanché cell.

The current from a simple copper-zinc cell soon gets weaker and weaker, because a layer of hydrogen collects on the copper plate and prevents the acid from reaching the copper, while the hydrogen tends to set up a current in the opposite direction. That is, the layer of hydrogen forms a *new pole* between the copper and the zinc, and the cell is said to be *polarized*.† Polarization in the simple voltaic cell is a very serious disadvantage, because any cell that is to be used for practical purposes must give a *steady* current. Hence all practical voltaic cells have special arrangements for reducing polarization, i.e. for *depolarizing* the cell. In most

practical voltaic cells an oxidizing agent is used (e.g. manganese dioxide or potassium dichromate—$K_2Cr_2O_7$) to oxidize the hydrogen on the positive plate to water and thus depolarize the cell.

THE LECLANCHÉ CELL

The disadvantages of (*a*) *polarization*, and (*b*) '*local action*' are overcome in the voltaic cell invented by Leclanché in 1868, and the *Leclanché cell*† is the only voltaic cell in everyday use today. The positive pole is made of *carbon* and the negative pole of *zinc*. The carbon rod or plate is surrounded by a coarse mixture of *manganese dioxide* (an oxidizing agent) and *carbon* (to increase the surface area of the positive pole and to absorb some of the hydrogen) contained in a porous pot or bag (see Fig. 185). The cell is filled with a saturated solution of *ammonium chloride*. When the poles are connected by a wire, a current flows and the zinc begins to dissolve:

$$Zn + 2NH_4Cl \longrightarrow ZnCl_2 + 2NH_3 + H_2$$

The grains of carbon absorb some of the hydrogen, which the manganese dioxide begins to oxidize to water:

$$H_2 + MnO_2 \longrightarrow H_2O + MnO \text{ (manganous oxide).}$$

If only a small current is being taken from the cell (e.g. when used as part of the 'high tension battery' of a wireless set) the manganese dioxide can oxidize the hydrogen as fast as it is formed, depolarizing the cell continuously and maintaining a steady current. If a bigger current is taken from the cell (e.g. when used for lighting a pocket-lamp bulb) the manganese dioxide cannot oxidize the hydrogen as fast as it is formed and the current soon falls off owing to polarization. After a rest, however, the manganese dioxide depolarizes the cell. Hence a Leclanché cell is only suitable for supplying small currents for short periods, so that the cell has time to depolarize before it is used again. It is a very useful cell, however, for working electric bells.

THE DRY CELL

The Leclanché cell is most commonly met with in its 'dry' form (see Fig. 186). The *zinc* container forms the negative pole. The central *carbon* rod forms the positive pole and this is surrounded by a *damp* depolarizing mixture of *ammonium chloride*, *manganese dioxide*, and *carbon*, contained in a porous bag. The space between this bag and the zinc container is filled with a *damp* paste of ammonium chloride and flour. These *dry cells* are made in millions each year for use with pocket-lamps, bells, telephones, and wireless sets, where only small currents are required at infrequent intervals.

THE SOURCE OF ENERGY IN VOLTAIC CELLS

By *the Principle of Conservation of Energy*, electrical energy can only be obtained by changing an equivalent amount of some other form of energy into electricity. In *voltaic cells* one of the metals (usually zinc) dissolves to produce electricity and a salt of the metal. This salt possesses less potential (chemical) energy than the metal from which it was formed, e.g. when 1 gm. of zinc dissolves to form zinc sulphate solution, about 1,300 calories of heat are set free.

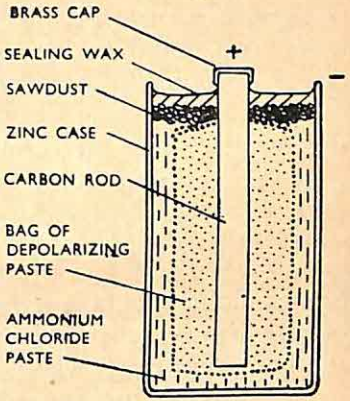


FIG. 186. 'Dry' cell.

In a simple copper-zinc cell this energy appears as electricity and not as heat. Hence the electricity is obtained at the expense of the zinc that dissolves.

This is a very expensive way of producing electricity, but as a 'dry' cell is self-contained and easy to carry about, it provides a very *convenient* way of producing small electric currents. It would be far too expensive, however, to light a house or drive an electric fan with electricity obtained from voltaic cells. It is much cheaper to produce large quantities of electricity from a *dynamo* driven by a heat-engine whose energy is obtained from coal or oil.

Voltaic cells, producing electrical energy directly from the chemical energy of a metal, are called *primary batteries*. When larger quantities of electricity are required out of reach of a wired supply from a dynamo (e.g. in a motor-car), an *accumulator* (or *storage cell*) is used. This type of cell is called a *secondary battery* because it does not produce electrical energy directly, but must first be charged with electricity, electrical energy being converted into chemical energy and stored in that form until an electric current is required. The stored-up chemical energy is then converted into electrical energy once more as the cell is discharged.

Most accumulators consist of two *lead* plates surrounded by dilute sulphuric acid (1 part of acid in 4 parts of water). When *charged*, one plate becomes covered with grey, spongy, metallic *lead* (*−ve*), and the other with brown *lead peroxide* (*+ve*). As the cell is *discharged* (i.e. when current is taken from it), both the metallic lead on the negative plate and the lead peroxide on the positive plate are converted into insoluble *lead sulphate*,† which covers both plates.

$$2PbSO_4 + 2H_2O \underset{\text{Charge}\rightarrow}{\overset{\leftarrow\text{Discharge}}{\rightleftharpoons}} Pb + PbO_2 + 2H_2SO_4$$

Some sulphuric acid is removed from the battery acid during discharge, hence the specific gravity of the acid becomes less, in fact the best way to tell whether a lead accumulator is fully charged or not is to measure the specific gravity of the battery acid with a hydrometer. When the cell is recharged by passing an electric current through it (from a dynamo) the lead sulphate is changed back again to metallic lead on the negative plate and to lead peroxide on the positive plate, sulphuric acid being set free at the same time, so that the specific gravity of the battery acid increases. When the battery is fully charged it 'gasses' freely, bubbles of hydrogen being set free at the negative plate and bubbles of oxygen at the positive plate.

The chief disadvantage of lead accumulators is their great

weight. They are also damaged by too rapid charging or discharging.

Set up a simple lead accumulator by fixing two lead plates to a strip of wood by two terminals† and dipping the plates in a jar containing dilute sulphuric acid (1 in 4). (*Care.* Battery acid is many times stronger than the dilute sulphuric acid in your bench reagent bottles.) Set up the circuit shown in Fig. 187, with a pocket-lamp

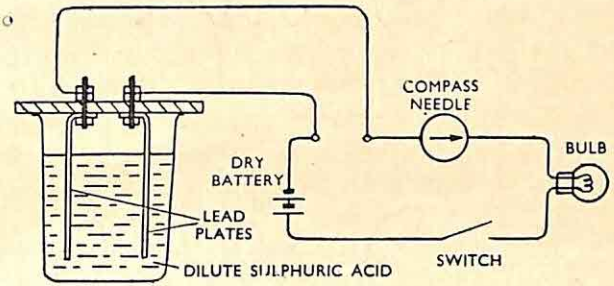


FIG. 187. Charging a simple lead accumulator.

battery (2 or 3 cells), a pocket-lamp bulb, a simple current-detector,[1] a tapping-key, and your simple lead accumulator in series. Hold down the tapping-key and notice that the bulb lights faintly and that the needle of your current-detector is deflected* to one side, showing that a small current is flowing through your 'accumulator'. Notice that bubbles of gas form on both plates and that after a few minutes one plate turns brown while the other remains grey. After 'charging' for a few minutes, replace the pocket-lamp battery with a piece of copper wire and 'discharge' your 'accumulator' by pressing the tapping-key.

Notice that the bulb lights brightly for a short time, showing that your 'accumulator' is discharging much more rapidly than you charged it. Notice also that the needle of your current-detector is deflected to the opposite side, showing that the current is flowing in the reverse direction.

[1] e.g. a pocket-compass lying over a wire or strip of metal.

THE SCREW RULE

The last experiment shows that we can use a compass-needle to show which way a current is flowing in a neighbouring wire. Learn the '*Screw Rule*': *Imagine a right-handed screw being screwed along the wire in the direction in which the current is flowing*

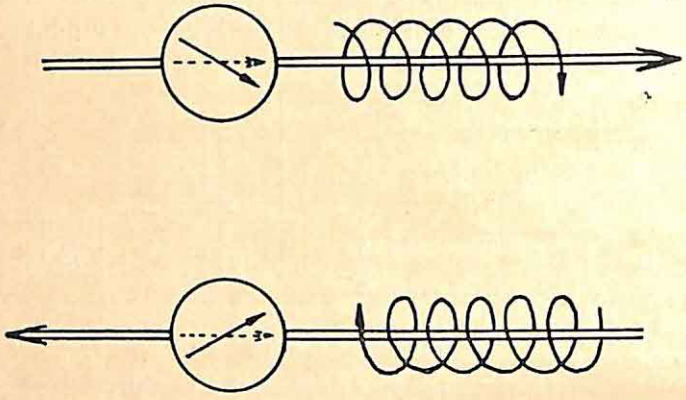


FIG. 188. The screw rule.

(copper to zinc with a voltaic cell). *Then the direction in which the screw turns gives the direction of the magnetic lines of force* (i.e the direction in which the N-pole of the compass-needle will be deflected).

Verify this rule with your 'charging-circuit' (see Fig. 187).

(See *General Science Workbook*, Book Two, pp. 109–11, for learning exercises on this chapter.)

Chapter XV

THE EFFECTS PRODUCED BY ELECTRIC CURRENTS

Electrical energy is easily converted into heat and into chemical energy. In Book Three, Chap. V, we saw that there is a magnetic field round a wire carrying an electric current. In everyday life we are mainly concerned with the *effects* produced by electric currents. These can be grouped under three headings: (i) *the heating effect*, (ii) *the chemical effect*, and (iii) *the magnetic effect*. We use the *heating effect* in electric heating and lighting: we use the *chemical effect* in electrolysis, electroplating, and secondary batteries; we use the *magnetic effect* in dynamos, motors, telephones, and most electrical measuring-instruments.

THE HEATING EFFECT OF AN ELECTRIC CURRENT

We have seen that one of the great practical advantages of electricity over other forms of energy is that it can travel long distances through *conductors*, usually copper wires. Whenever an electric current passes through a wire, however, some electrical energy is converted into heat-energy. When small currents are passed through thick copper wires, the rise of temperature is only small and the heat produced soon escapes into the surrounding air; but when large currents are passed through thin wires of high '*resistance*', much heat is produced and the wire gets very hot (as we have seen by the expansion produced in a nichrome resistance wire, Book Three, Chap. I). By passing measured electric currents through coils of fine resistance wire surrounded by water in a calorimeter, it can be shown that *the heat produced is proportional to* (*a*) *the square of the current*,[1] (*b*) *the resistance*[1] *of the wire*, and (*c*) *the time for which the current flows.*

[1] We shall learn in later lessons how *currents* and *resistances* are measured.

ELECTRIC HEATERS

The heating effect of an electric current is widely used in every-day life to provide a clean and convenient source of heat. Electric heaters, cookers, irons, etc., get their heat from the heating effect of an electric current passing through coils of resistance wire. The wire used for this purpose is usually *nichrome*, an alloy of nickel, chromium, and iron, which has a resistance 50 times greater than that of copper and does not oxidize and burn away at a red heat. Nichrome wire is often flattened out into ribbon, and this resistance wire or ribbon is wound on supports made of fire-clay, mica, or some other non-conductor that will stand high temperatures.

Power-stations usually supply the same electrical energy at a cheaper rate for heating than for lighting purposes, because electric heaters and cookers are used mainly in the day-time when very little current is being used for lighting. By reducing the rate for electric power used for heating purposes, a power-station gets a steady load for its dynamos during the day and not merely a heavy load in the evening when everyone switches on the electric light. A steady output results in cheaper running expenses for the power-station per unit of current produced.

FUSES

A *fuse*† is a short length of thin wire of low melting-point, introduced into an electric light or 'power' circuit to protect the rest of the wiring from a dangerous 'overload'. If the current becomes too great for safety, the fuse-wire gets hot, melts, and stops the flow of current before the low-resistance wires in the rest of the circuit become dangerously hot. The fuse is usually enclosed in a fire-proof porcelain holder or glass tube. In this way, the damage done by an 'overload' or 'short-circuit' is confined to a safe place where it is easily repaired. If the new piece of fuse-wire 'blows' when the current is switched on again, it is important to find the cause of the trouble—probably a 'short-circuit' (i.e. a direct connexion between two opposite wires of the circuit).

In an electric heater the resistance wire seldom becomes more than red-hot. At a higher temperature the metal would combine with the oxygen of the air and burn away. If a wire is to be heated white-hot so that it gives out light (as well as heat), it must be

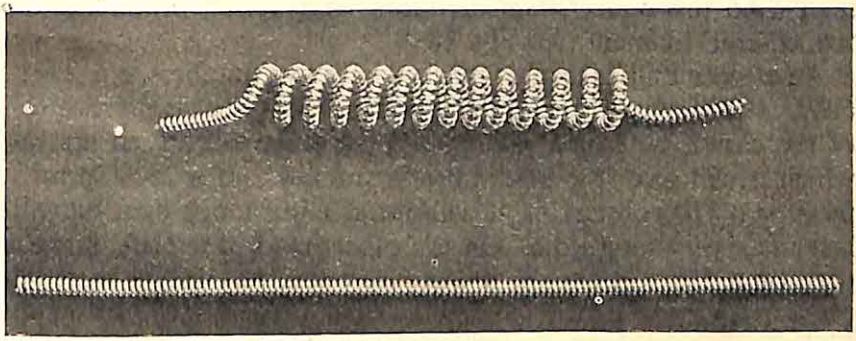


FIG. 189. Coiled-coil filament lamp. Pieces of the Mazda coiled-coil filament (above) and the filament coil in an ordinary gas-filled lamp (below).

heated in the absence of oxygen. The earliest electric bulbs used a thin thread (or *filament*) of *carbon*, enclosed in a glass bulb from which all the air had been removed. These *carbon filament bulbs* gave a yellowish light because the filament could not be heated above about 1,750° C., and they converted only about 0·5 per cent. of the input of electrical energy into light-energy, i.e. over 99 per cent. was converted into heat; in fact, carbon filament bulbs are used today for heating rather than for lighting.

The filaments of modern electric bulbs are made of *tungsten*,† a metal that melts above 3,500° C. The first tungsten filament lamps had vacuum bulbs, and the filament could be heated to about 2,500° C., giving an efficiency of about 1 per cent., i.e. about twice the efficiency of a carbon filament bulb. At higher temperatures, however, the tungsten gradually vaporized and condensed on the inside of the bulb, blackening it and reducing the output of light. Modern electric lamps are 'gas-filled', the bulb being filled

with *argon* to slow down the vaporization of the tungsten. In such 'gas-filled' bulbs, the filament can be run at about 3,000° C., giving about twice as much light for the same electrical power. A later improvement was to wind the tungsten filament into a spiral coil so that the neighbouring coils 'keep each other warm', and a still later improvement was to first make a spiral of very tiny coils and then wind this coiled filament into a large spiral, producing the 'coiled-coil' filament (see Fig. 189).

An electric bulb is usually marked with two numbers, e.g. 230V, 60W. The first number shows the *voltage*† (or 'electrical pressure') of the circuit on which the lamp is designed to be used and on which it will give 'the best value for money'. If a 230V lamp is used on a 110 volt circuit, it will not give out much light but will last for a very long time. On the other hand, if a 110V lamp is used on a 230 volt circuit, it will give out a very bright light but only for a very short time, because the filament gets too hot on the higher voltage. The second number on the bulb shows the number of *watts* used, i.e. the *electrical power* required to light the lamp at the given voltage. Modern electric bulbs use from 0·5 to 1·0 *watts per 'candle-power'*, which means that only about 2 per cent. of the electrical energy is converted into light-energy. The remaining 98 per cent. is radiated as heat and wasted. It is interesting to compare the lighting efficiency of our best electric bulbs (about 2 per cent.) with that of the fire-fly (nearly 100 per cent.).

Electric discharge lamps, now used in many towns for street-lighting, work on an entirely different principle, an electric discharge being passed through mercury vapour or sodium vapour. Mercury vapour lamps give a greenish-blue light while sodium lamps give a golden-yellow light. These discharge lamps are several times more efficient than any filament lamps, but as they do not produce white light they are unsuitable for house-lighting. Another type of discharge lamp produces light by passing an electric discharge through a gas under a low pressure. *Neon* discharge tubes, for example, give a bright red light and are often bent into complicated shapes for use in advertising signs.

Modern tubular *fluorescent lamps*† are coated with '*phosphors*',†[1] which are chemical compounds that give out visible light when struck with invisible *ultra-violet rays*. The lamp-tube contains small quantities of *argon* and *mercury vapour* and two electrodes heated by an electric current. When the electric current is switched on, electrons pass from one electrode to the other and the argon begins to glow. This makes the mercury glow in turn, and if the tube were clear glass (uncoated with phosphors) all you would see would be a faint blue light from the glowing gas-mixture. This glow is very rich in *ultra-violet rays* and when these strike the solid phosphors coating the inside of the tube, the phosphors convert invisible ultra-violet light into visible light (with very little heat). The colour of the visible light given out is determined by the kind of phosphor used to coat the fluorescent tube. (The fire-fly produces 'cool light' by a similar principle, using energy from an enzyme-assisted chemical reaction to 'excite' a phosphor.) Fluorescent tubes are very efficient, and a 15 watt fluorescent lamp will give out as much light as a filament bulb using 40–60 watts.

THE ARC LAMP

If two *carbon rods* are allowed to touch each other and a sufficient current is passed, the ends of the rods become red-hot. If the carbon rods are now drawn 2–3 mm. apart, a blinding white *electric arc* is formed at the gap. This arc consists of white-hot carbon vapour at a temperature of nearly 4,000° C., the highest temperature that Man can produce. An arc lamp converts into light energy about 10 per cent. of the electrical energy supplied to it and provides our most efficient source of artificial light, but it has the disadvantage that the light is of an unpleasant colour. Arc lamps are used mainly for cinema-projectors and for search-lights, where it is important that the source of light should be concentrated at one focal point.

The high temperature of the electric arc is made use of in *arc welding*. An electric arc is produced at the point where two pieces

[1] Not, however, containing the element phosphorus, in spite of their name.

of metal are to be joined, and the high temperature soon melts a little of the metal and welds the two pieces together.

THE CHEMICAL EFFECTS OF ELECTRIC CURRENTS

In Book One, Chap. V, we saw that water containing a little sulphuric acid will conduct an electric current. This kind of conduction, however, is different from the conduction of a current by a copper wire, where the conductor remains unchanged after the passage of the current. The acidified water is split up into hydrogen and oxygen when a current passes through it. This decomposition of a substance by electricity is called *electrolysis*. Substances that conduct electric currents and at the same time undergo decomposition are called *electrolytes*.† Water solutions of practically all *acids*, *bases*, and *salts* are electrolytes. *Pure* water and most organic liquids, e.g. alcohol and oils, are *non-electrolytes*. During *electrolysis*, the current enters and leaves the electrolyte through conductors (usually metal plates) called *electrodes*. The electrode at which the current enters is called the *anode* (+*ve*); and the other, at which the current leaves, is called the *cathode* (−*ve*). During electrolysis, *hydrogen and the metallic elements are set free at the cathode, and all other elements at the anode.*

IONS

When inorganic acids, bases, and salts dissolve in water their molecules split up into smaller particles called *ions*.† For example, when sodium chloride dissolves in water, each molecule splits into a *sodium ion* and a *chlorine ion*, each ion having an electrical charge that gives the particle some special properties. A sodium ion has a positive charge and is represented by Na^+, while a chlorine ion has a negative charge and is represented by Cl^-. So the *ionization*† (or *dissociation*†) of sodium chloride in water can be represented by $NaCl \longrightarrow Na^+ + Cl^-$.

It is because of these electrical charges that electrolytes conduct electric currents, e.g. when sodium chloride solution is electrolysed, the positively charged sodium ions are attracted to the

negative electrode (or *cathode*) and the negatively charged chlorine ions to the positive electrode (or *anode*).

When we defined acids, bases, and salts (in Book Two, Chap. III), we could give no simple, yet complete, definition, but we can now define these three main classes of chemical compounds in terms of the *ions* that they form when they dissolve in water. *An acid is a substance that forms hydrogen ions* as the only positive ions in

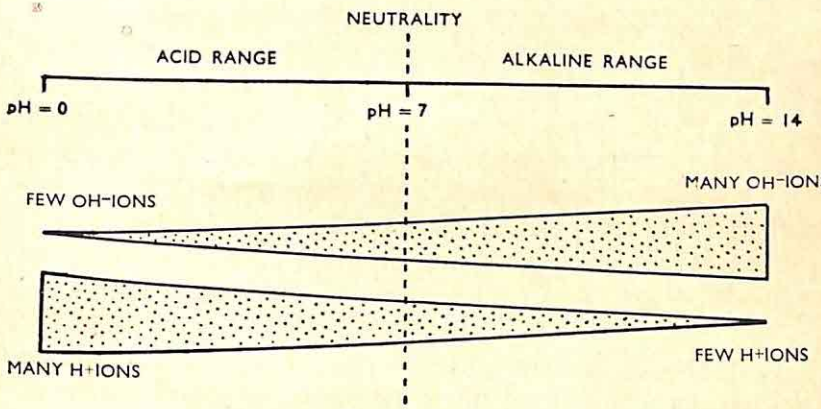


FIG. 190. The pH scale for measuring acidity and alkalinity.

solution, e.g. $HCl \longrightarrow H^+ + Cl^-$. *An alkali* (or soluble base) *is a substance that forms hydroxyl† ions* (negatively charged combinations of oxygen and hydrogen atoms) as the only negative ions in solution, e.g. $NaOH \longrightarrow Na^+ + OH^-$. *A salt is a substance that forms neither hydrogen nor hydroxyl ions* (but its solution contains *positive ions other than hydrogen ions* and *negative ions other than hydroxyl ions*), e.g. $NaCl \longrightarrow Na^+ + Cl^-$.

The degree of acidity or alkalinity of a solution depends on the extent to which it is split up into ions when dissolved in water, i.e. on the number of hydrogen ions or hydroxyl ions that the solution contains. For practical purposes, this is expressed by numbers on the pH *scale* (see Fig. 190). For example, pH = 0 represents maximum acidity (i.e. many H^+ ions and very few OH^- ions), pH = 14 represents maximum alkalinity (i.e. many

OH^- ions and very few H^+ ions), while pH = 7 represents chemical neutrality (i.e. equal numbers of H^+ ions and OH^- ions in the solution).

ELECTROLYSIS OF COPPER SULPHATE

Set up the circuit shown in Fig. 191. X is a pocket-lamp battery and Y a pocket-lamp bulb.

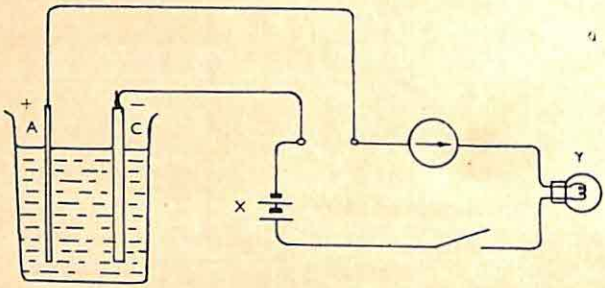


FIG. 191. Electroplating.

The anode A is a strip of thin copper sheet and the cathode C is a carbon rod[1] dipping into a 10 per cent. solution of copper sulphate in water acidified with dilute sulphuric acid. After the current has been flowing for 5 minutes, take out the electrodes and notice that the part of the carbon cathode that was beneath the surface is now covered with copper (or *copper-plated*). Reverse the battery connexions (i.e. make the copper-plated carbon the anode and the copper strip the cathode) and again pass the current for 5 minutes. Notice that the coating of copper disappears from the carbon. It is clear that *the electric current removes copper from the anode and deposits it on the cathode.*

ELECTRO-PLATING

In the last experiment we copper-plated the carbon cathode. If we use a nickel anode instead of copper and a solution of a nickel salt as the electrolyte, we can *nickel-plate* the cathode. This pro-

[1] e.g. the positive pole from an old pocket-lamp battery.

cess of covering a surface with a thin layer of metal, by electrolysis, is called *electro-plating* and it has many everyday applications. Gold, silver, and chromium can be deposited on other metals in the same way, but in actual practice it is usual to apply a layer of copper before plating with any of these other metals.

ELECTROTYPES

Copper-plating is used in the printing industry for *electrotyping*,† i.e. producing a number of printing-plates from a single page of type. When a page of type has been set up, it is pressed into a sheet of wax. This wax mould is covered with a thin layer of graphite to form a conducting surface, and is then made the cathode in a solution of copper sulphate with a copper anode. When a current is passed, a layer of copper is deposited on the wax mould, giving an exact copy of the original type. The wax is then melted away from the copper and the thin sheet is strengthened by pouring melted type-metal over its back. The copper *electrotype* is then used for printing, and the type distributed for use in another book. Electrotypes are much lighter than lead type and take less room to store. By making several electrotypes of each page, a large number of copies can be printed much more rapidly than by using only the original single set of type for the book.

The same principle is used in making gramophone-records. The original record is traced by the vibrating needle on a disc of wax. This wax record is then dusted with graphite and electro-plated with copper. The thin sheet of copper, with ridges corresponding to the wavy grooves in the original wax record, is backed with metal and pressed on flat discs of a black, plastic material that has been softened by warming, thus 'printing' a copy of the original wax sound-track on the record we play on a gramophone.

ELECTRO-METALLURGY

Several common metals are purified by electrolysis, e.g. the copper that is deposited on the cathode of an electrolytic cell is chemically pure. Such pure copper is in great demand for electrical

purposes owing to its low resistance. Hence, copper used for electric wiring is refined by electrolysis. Thick sheets of crude copper are hung in a solution of copper sulphate and form the anodes. The cathodes are thin sheets of pure copper, and, when the current flows, copper is removed from the crude anodes and pure copper is deposited on the cathodes. Impurities fall to the bottom of the liquid.

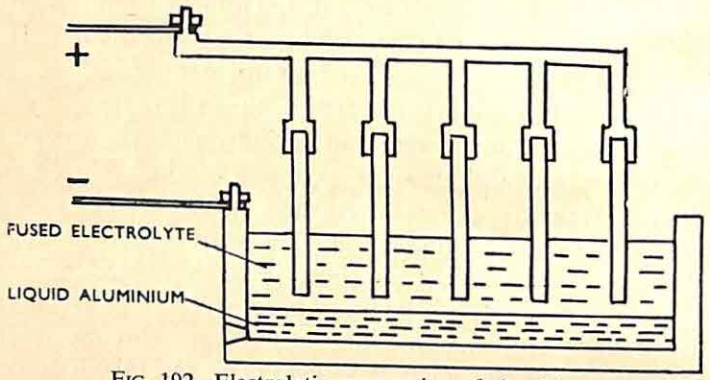


FIG. 192. Electrolytic preparation of aluminium.

Gold and silver can be refined by electrolysis in the same way.

Aluminium is now produced in enormous quantities by electrolytic methods. Although aluminium is one of the commonest elements in the Earth's crust (7 per cent.), only in recent years has it become possible to extract the metal from its ores, because of the difficulty of reducing its oxide. Carbon, for example, will not reduce aluminium oxide to metallic aluminium. *Bauxite* (aluminium oxide with some loosely-combined water) is the ore used as a source of aluminium. It is first purified and then heated to convert it into the anhydrous oxide, Al_2O_3, which is dissolved in fused *cryolite*† (a fluoride of aluminium and sodium). This liquid mixture (at a temperature of 800–1000° C.) is electrolysed in an iron vessel lined with carbon (which serves as the cathode) with large carbon anodes dipping into the mixture (see Fig. 192). The liquid aluminium collects on the negative carbon floor and oxygen

is set free at the carbon anodes, which gradually burn away. The cryolite is not used up during electrolysis and it is only necessary to add more aluminium oxide from time to time.

Magnesium is produced in a similar way by the electrolysis of fused magnesium chloride. Sea-water is now used as a source of this magnesium chloride.

In Chap. II we have referred to the commercial importance of the electrolysis of brine (saturated sodium chloride solution) in the manufacture of sodium hydroxide. Metallic sodium and potassium are also prepared by electrolysis.

QUANTITY OF ELECTRICITY

As Faraday showed by experiment in 1832, when solutions of copper sulphate or silver nitrate are electrolysed, the weight of copper or silver deposited depends only on *the quantity of electricity* that passes. This gives us a direct method of measuring *quantity of electricity*, and the practical unit of quantity of electricity is called the *coulomb*† (after the French scientist of that name). *A coulomb is the quantity of electricity that will deposit* 0·001118[1] *gm. of silver.*

The 'strength' of an electric current is *the rate of flow of electricity* through the circuit, i.e. the quantity of electricity passing per second. The practical *unit of current* is the *ampere*† (so called after the French scientist of that name). *An ampere is a flow of one coulomb per second,* or, *an ampere is a current that will deposit* 0·001118[1] *gm. of silver in one second.*

We can test the accuracy of an *ammeter*† (or ampere-meter), therefore, by comparing the reading shown by the instrument with the weight of silver deposited during a known time from a solution of silver nitrate in an electrolytic cell placed in series with the ammeter.

The following everyday examples will give you a better idea of this unit of current: (i) An ordinary 60-watt bulb takes a current of about one-quarter of an ampere from the 230-volt house supply;

[1] You need not memorize this figure.

(ii) a motor-car battery is charged at a rate of about 8–10 amps. when the car is running in the day-time; (iii) a 36-watt motor-car head-lamp bulb takes 3 amps. from a 12-volt accumulator; (iv) the 'self-starter' motor of a car may take as much as 100 amps.

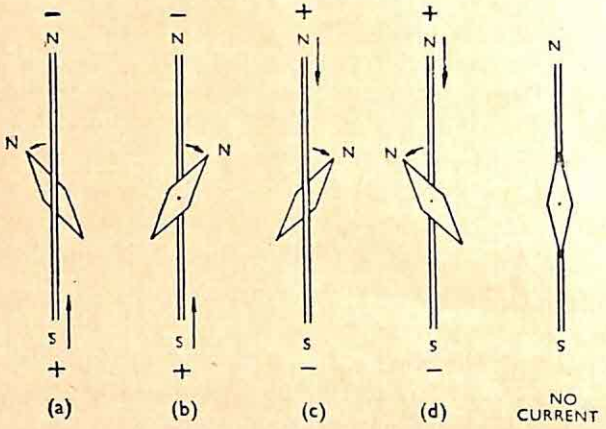


FIG. 193. Oerstedt's experiments.

from the accumulator (hence it is very bad for the car-battery if the starter-switch is kept closed for more than a second or two); (v) a 'valve' in a wireless receiver takes only a few milli-amperes (i.e. thousandths of an ampere) from the 'high tension' supply; (vi) the currents flowing in long-distance telephone-wires are only small fractions of a milli-ampere; (vii) fuses in house-lighting circuits usually melt when the current reaches 5 amps.

THE MAGNETIC EFFECTS OF ELECTRIC CURRENTS

When a wire is held parallel to a compass-needle and an electric current is passed through it, the needle is deflected to one side. The Earth's magnetic field tends to turn the compass-needle into a north and south direction, while the action of the current tends to turn the needle into an east and west direction.

Under the combined influence of these two turning forces, at

right angles to each other, the needle sets itself in such a direction that the two forces balance.

This magnetic effect of an electric current was first noticed in 1820 by Oerstedt, a Danish scientist, who thus discovered the close relation between *magnetism* and *electricity*: *that there is always a magnetic field round a conductor carrying an electric current.*

Experiment:

(*a*) Hold the wire carrying current *over* the compass needle. Notice the direction in which the needle is deflected.

(*b*) Hold the wire carrying current *under* the compass-needle. Notice the direction in which the needle is deflected.

(*c*) Reverse the current in the wire; hold the wire *over* the compass-needle. Notice the deflection.

(*d*) Repeat with the wire *under* the needle. The four results are shown in Fig. 193.

These results are summarized in the Swimming Rule: '*Suppose a man is swimming with the current, and with his face towards the compass-needle, then the N-pole is deflected towards his left hand.*'

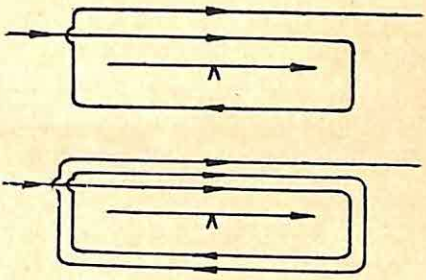


FIG. 194. Simple current detector.

If the wire carrying the current is held first *over* and then *under* the compass-needle, the needle is deflected in opposite directions, but if the wire is carried first above the needle and then bent back again below the needle, as shown in Fig. 194, we get twice the turning effect. If we take the wire several times round the needle, over and under in turn, we can multiply the turning effect as much as we like. Such an arrangement gives us a very delicate method for detecting electric currents and for comparing their strengths, since the stronger current produces the greater deflection of the needle. This principle is applied in one type of current-measuring instrument (see p. 364).

THE MAGNETIC FIELD ROUND A STRAIGHT CONDUCTOR

If a thick copper wire is fixed vertically through a small hole in a horizontal sheet of cardboard covered with iron filings and a strong current (of at least 10 amps.) is then passed through the vertical wire, the filings arrange themselves in circles round the wire. With the current flowing downwards, pocket-compasses placed on the cardboard show that the direction of the lines of force is as shown in Fig. 195 (as you would expect from the Swimming Rule). When the current is reversed, the direction of the lines of force is also reversed.

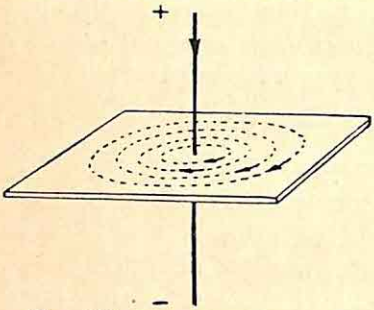


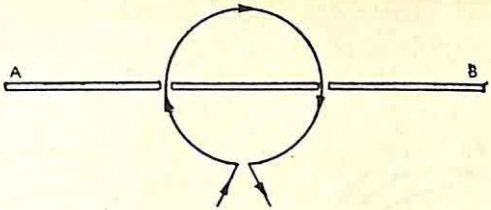
FIG. 195. Magnetic field round straight conductor.

THE MAGNETIC FIELD ROUND A CIRCULAR CONDUCTOR

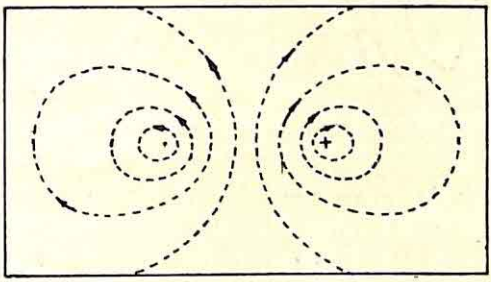
If a vertical *ring-coil* of wire is used instead of a single vertical wire, and iron filings are spread on the horizontal sheet of cardboard (AB in Fig. 196), when the current is passed the lines of force are found to be arranged as in Fig. 196. This magnetic field is of the same type as that surrounding a very short bar-magnet or a magnetic disc. If the number of circular turns of wire in the ring-coil is increased, the magnetic field will be strengthened in the same proportion.

THE MAGNETIC FIELD WITHIN AND AROUND A SOLENOID†

If the wire is wound into a close spiral to form a cylindrical coil (or *solenoid*), the magnetic field produced on passing a current has its lines of force arranged as shown in Fig. 197. Notice that these lines of force are just like those round a long bar-magnet. If a soft-iron rod is placed inside such a cylindrical coil the soft iron becomes a strong magnet *while the current is flowing*. If a steel rod is placed inside the coil it becomes a permanent magnet; and this is how permanent magnets are usually made.



HORIZONTAL SECTION THROUGH COIL



VERTICAL SECTION THROUGH COIL

FIG. 196. Magnetic field round ring-coil.

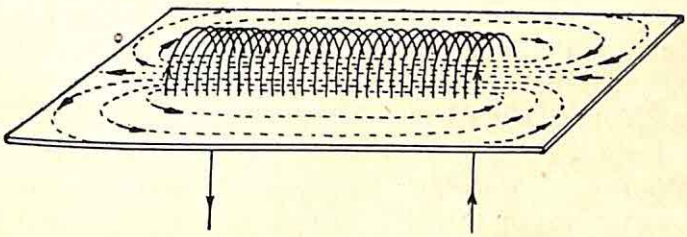
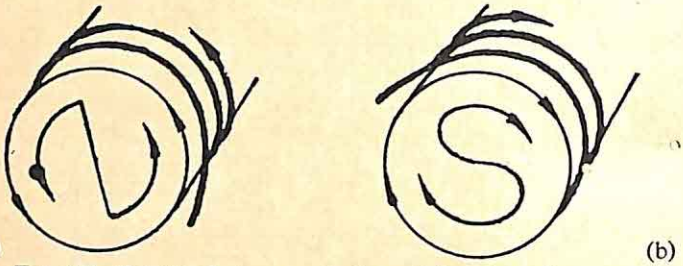


FIG. 197. Magnetic field round solenoid.

The solenoid behaves like a bar-magnet while the current is flowing and its ends have polarity that can be tested by using a compass-needle.

The ends of the solenoid that correspond to N-pole and S-pole are easily remembered from the diagrams shown in Fig. 198. Notice the directions of all five arrows in each diagram.

It is sometimes necessary to find (*a*) which end of a coil corresponds to the N-pole of a magnet, or (*b*) which way the current is flowing in the coil. The easiest way of doing this is to remember the diagrams in Fig. 198. If the current flows as in (*a*), that end of



(a) (b)

FIG. 198. Relation between polarity and direction of current in a solenoid.

the coil is a N-pole (or conversely, if that end of the coil is found to be a N-pole then the current is flowing *anti-clockwise* as we face that end of the coil). If the current flows as in (*b*), that end of the coil is a S-pole (or conversely, if that end of the coil is found to be a S-pole, then the current is flowing *clockwise* as we face that end of the coil).

Another way of remembering polarity is by learning *the 'grip' rule*. Imagine that you grip the solenoid with the fingers of your right hand parallel to the wires and pointing in the direction of the current. Your extended thumb then points to the N-pole of the solenoid.

(See *General Science Workbook*, Book Two, pp. 111–34, for learning exercises on this chapter.)

CHAPTER XVI

THE MEASUREMENT OF ELECTRICITY

CURRENT-MEASURING INSTRUMENTS

For everyday practical purposes, electric currents are nearly always measured by using their magnetic effect, and there are three main

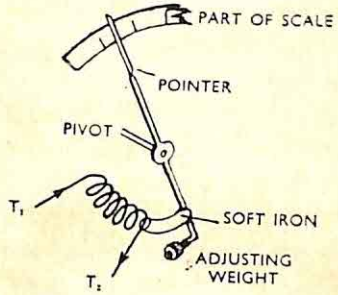


FIG. 199. Moving-iron (attraction) instrument (after Pearce).

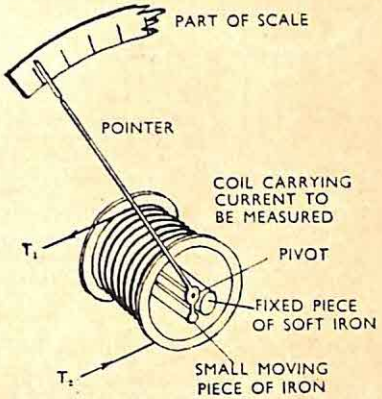


FIG. 200. Moving-iron (repulsion) instrument (after Humby).

types of magnetic current-measuring instruments: (i) *moving-iron instruments*, (ii) *moving-magnet instruments*, and (iii) *moving-coil instruments*.

MOVING-IRON INSTRUMENTS

The simplest and cheapest kind of magnetic current-measuring instrument is the *moving-iron* instrument, which has the further advantage that it will measure both *direct currents* (D.C.) and *alternating currents* (A.C.). In the type shown in Fig. 199, the current to be measured passes through a coil between the two terminals T_1 and T_2, drawing a piece of soft iron into the coil and thus moving the pointer across the scale. The attraction of

the soft iron and the movement of the pointer depend upon the strength of the current passing through the coil.

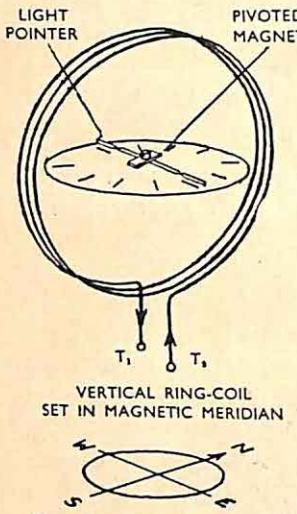


FIG. 201. Moving magnet instrument.

In the type shown in Fig. 200, two pieces of soft iron (one fixed and the other pivoted to a pointer) lie side by side and are surrounded by the coil carrying the current to be measured. Both pieces of iron are magnetized when a current passes through the coil, and as their like poles are together they repel each other, moving the pointer across the scale. The repulsion between the two pieces of iron and the movement of the pointer depend on the strength of the current passing through the coil.

The pointer-scale of a moving-iron instrument is graduated by passing currents of different strengths through the instrument when connected in series with an accurate standard instrument.

MOVING-MAGNET INSTRUMENTS

The earliest current-measuring instruments were of the *moving-magnet* type, but although they are cheap and easy to make they are never used for everyday electrical measurements outside a school laboratory.

The instrument is shown in Fig. 201. The short magnet lies at the centre of the vertical ring-coil and is attached to a long light pointer moving over the circular scale. Before any current is passed the instrument must be placed so that coil and magnet are parallel (i.e. so that they lie in the magnetic meridian). This is the main practical disadvantage of the instrument. When the current flows, the needle is deflected and will come to rest in a position that is determined by the strength of the current. The instrument cannot be used near other magnets, or near magnetic materials.

Most modern current-measuring instruments are of the *moving-coil* type, which is free from the disadvantages of the moving-magnet and moving-iron types. In this kind of instrument, there is a large, strong, *fixed magnet*, and a small, light, *moving coil* through which the current flows.

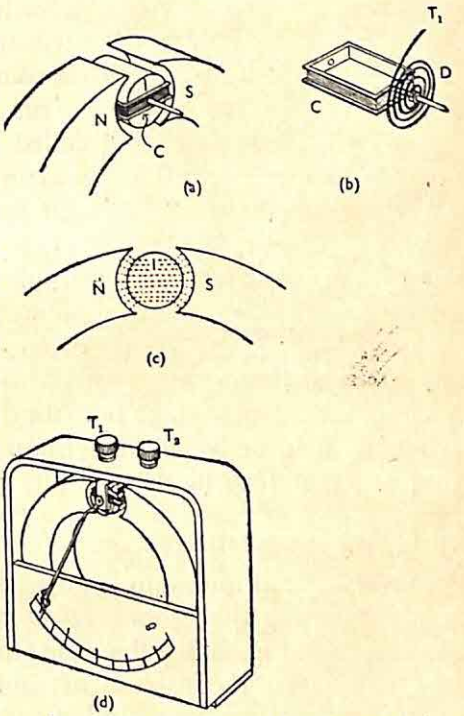


FIG. 202. Moving-coil instrument (after Calvert).

We have seen (p. 360) that when a current flows through a ring-coil of wire, the coil behaves like a very short bar-magnet or magnetic disc. If such a coil is hung between the poles of a circular magnet, it will tend to turn so that the N-pole side of the coil faces the S-pole of the permanent magnet.

As shown in Fig. 202, a cylindrical soft-iron core I is usually placed inside the coil so as to concentrate the magnetic lines of force between the poles NS of the permanent magnet. A rectangular coil C, of many turns of thin, insulated wire, is free to turn round this soft-iron core, and its two ends are connected by two spiral springs to the terminals of the instrument (D is one of these springs).

When no current is passing, the springs hold the pointer at the zero mark on the scale, i.e. with the plane of the coil parallel to the magnetic lines of force crossing the gap between the poles of the permanent magnet. But for the springs, when a current flows round the coil it would turn until its N-pole side faced the S-pole

of the magnet, but the spring checks the coil before it reaches this position. As the coil turns it twists the spring, and the larger the current the more the moving coil can twist the spring.

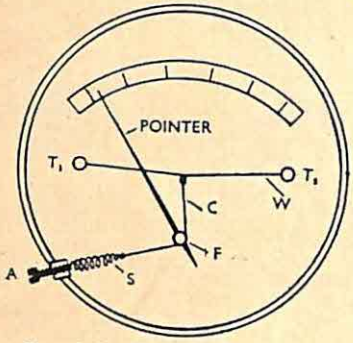


FIG. 203. Hot-wire instrument (after Humby).

Such an instrument is used for measuring very small currents (e.g. thousandths of an ampere). For larger currents a similar instrument called an *ammeter* is used, with its scale marked in *amperes*. A moving-coil instrument is made so that the turning effect on the coil is directly proportional to the current, hence the pointer-scale can be marked in equal divisions.

Moving-coil instruments have the great practical advantage that they are not affected by the Earth's magnetic field or by any neighbouring magnets or pieces of iron, and can therefore be used in any position.

HOT-WIRE INSTRUMENTS

A very cheap and simple current-measuring instrument can be made by using the *heating effect* of the current to be measured. As shown in Fig. 203, a fine resistance-wire W is stretched between the terminals of the instrument; and when a current passes through it, its temperature rises and it expands in length. The spring S pulls the thread C, which is wrapped round the axle F of the pointer, thus moving the pointer across the scale. A hot-wire instrument will measure both D.C. and A.C.

ELECTRICAL UNITS

In our lessons on heat we found that for heat to pass from one body to another, a difference of *temperature* is necessary, i.e. heat flows from the body at the higher temperature to the body at the lower temperature.

In the same way, in order that water may flow along a pipe,

there must be a difference between the *pressures* at the two ends of the pipe. In much the same way, in order that an electric current may flow along a wire, there must be a difference of '*electrical pressure*' at the two ends of the wire.

ELECTROMOTIVE FORCE—POTENTIAL DIFFERENCE†

The scientific name for this 'electrical pressure' is *electric potential*, and if one point in a circuit is at *a higher potential* than another, then electricity will flow from the first point to the second. The *potential difference* between two points is measured in *volts*,† i.e. *the volt is the unit of electric potential* (just as the ampere is the unit of electric current).

If we wanted to describe the flow of water through a pipe we could state the *quantity* of water flowing per second (i.e. the rate of flow) and also the '*head*' of water (i.e. its pressure or 'driving-power'). In the same way, we measure the *rate of flow* of an electric current in *amperes* and the '*electric pressure*' in *volts*. For example, a single-cell lead accumulator has a *potential difference* between its terminals of about 2 volts, or it is said to have an *electromotive*† *force* of 2 volts for driving a current along a wire connecting the two terminals. In a Leclanché cell, or a 'dry' cell, the chemical action inside the cell produces an electromotive force (E.M.F.) of 1·5 volts. A simple copper-zinc cell has an E.M.F. of just over 1 volt.

ELECTRIC POWER

The main advantage of electricity over other forms of energy is that electric currents can be made to travel for long distances through wires, thus *carrying energy* to a distant point where it can be made to *do work*. We have learnt that *power* is the *rate of doing work*, and that mechanical energy is measured in *horse-power* units (33,000 foot-pounds per minute). Electrical energy is measured in *watts*. There is a simple relation between the units of *electric power*, *current*, and *potential*: *If a current of one ampere flows at an E.M.F. of one volt, then the power is one watt*, or

Current (*in amperes*) × E.M.F. (*in volts*) = power (*in watts*), or

Amperes × volts = watts.

One Kilowatt = 1,000 watts. One H.P. = 746 watts. Therefore one Kilowatt = about $1\frac{1}{3}$ H.P.

Electrical energy is measured (and sold) in '*Board of Trade Units*', usually called simply '*units*', and this practical unit of *work* is the '*Kilowatt-hour*'. Thus, for example, when a current of 1 amp. at an E.M.F. of 100 volts flows for 10 hours it supplies one Kilowatt-hour of electrical energy.

To understand the relation between amperes, volts, and watts, you should work out some numerical examples.

ELECTRICAL RESISTANCE

We now know something about two of the quantities involved in the flow of a current along a conductor: *current* (measured in amperes) and *potential difference* (measured in volts). The third quantity is the *resistance of the conductor.*

Materials that allow a current to pass through them are called *conductors* (e.g. all metals and carbon), and their electrical *resistance* is said to be *low.* Materials that do not allow a current to pass are called *non-conductors* (e.g. glass, porcelain, rubber, paper, cotton, silk) and their electrical *resistance* is said to be very *high.* Some conductors allow a current to flow more readily than others, hence it is often necessary to *measure* the *resistance* of materials.

The *unit of resistance* is called an *ohm*,† after the German scientist of that name who, in 1827, carried out experiments on the flow of electric currents through metallic conductors, and found that *the current* (I) *flowing in a conductor is exactly proportional to the difference of potential* (V) *between its ends*, or $\frac{V}{I} = a$ *constant* (R). This is known as *Ohm's Law of Electrical Conduction.*

Hence every conductor has its own value for the ratio $\frac{V}{I}$ and this electrical constant R is called its *resistance.* When the current I is measured in *amperes* and the *potential difference* V in *volts*,

then the resistance R is in *ohms*; e.g. if *one volt* applied to the ends of a wire causes *one ampere* to flow through it, then the wire has a resistance of *one ohm*. For practical purposes, the *ohm* is defined as *the resistance of a column of pure mercury at* 0° *C.*, 106·3 *cm. long and* 1 *sq. mm. in cross-section.*[1]

We have now defined two of the international electrical units: the *ampere* (p. 357) and the *ohm*, so we can now define the *volt* (by Ohm's Law) as *that potential difference or electromotive force which drives a current of one ampere through a conductor whose resistance is one ohm.*

Ohm's Law is of the greatest importance in nearly all electrical measurements and calculations. To take a very simple example: If a motor-car head-lamp bulb is marked 12V 36W we know that it takes a current of 3 amps. and we can find the resistance of its filament, when hot, from $R = \frac{V}{I} = \frac{12}{3} = 4$ ohms.

You should apply Ohm's Law in numerical calculations of

Current: $I \text{ in amperes} = \frac{V \text{ in volts}}{R \text{ in ohms}}$;

E.M.F.: $V \text{ in volts} = I \text{ in amperes} \times R \text{ in ohms}$;

Resistance: $R \text{ in ohms} = \frac{V \text{ in volts}}{I \text{ in amperes}}$.

AMMETERS AND VOLTMETERS

A moving-coil current-measuring instrument like that described on pages 364–6 can be used either as an *ammeter* for measuring *current* in *amperes* (Fig. 205), or as a *voltmeter*† for measuring *E.M.F.* in *volts* (Fig. 204), by connecting known resistances with the instrument. Suppose that the moving coil has a resistance of 5 ohms and that a current of 0·015 amp. moves the pointer right across the pointer-scale. The scale could be divided into 15 divisions, each of which would be 0·001 amp., i.e. one milli-ampere, and the instrument would be a *milli-ammeter*. By Ohm's Law

[1] There is no need to remember these figures.

($V = IR$), when 0·015 amp. flows through a resistance of 5 ohms, the E.M.F. is $0{\cdot}015 \times 5 = 0{\cdot}075$ volts, so that each division of the scale would represent 0·005 volts, i.e. 5 *milli-volts*, and the same instrument can also be used as a *milli-voltmeter*.

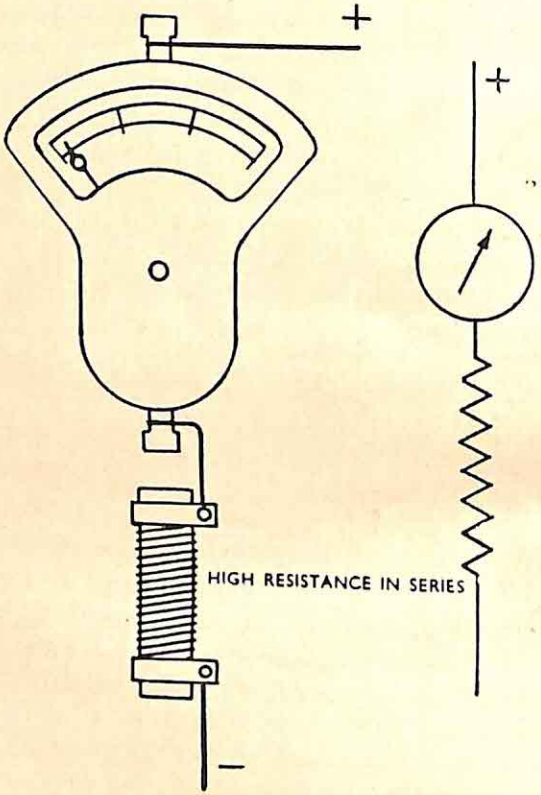


FIG. 204. Voltmeter.

Now suppose we want to measure volts instead of milli-volts. If a resistance of 495 ohms is connected *in series* with the instrument, as shown in Fig. 204, the current will have to pass through a total resistance of $495+5 = 500$ ohms and it will need an E.M.F. of $0{\cdot}015 \times 500 = 7{\cdot}5$ volts to drive the same *current* through the instrument and produce a full-scale deflection. Hence the instru-

ment becomes a *voltmeter*, reading 0–7·5 volts, instead of a milli-voltmeter reading 0–75 milli-volts. Similarly, by putting a 9995 ohm resistance in series with the 5 ohm moving coil, the same instrument can be used to measure any E.M.F. between 0 and 150 volts. Notice that the same *maximum current* of 0·015 amp. passes through the moving coil in every case.

If the same instrument is to be used for measuring 0–15 *amps.*

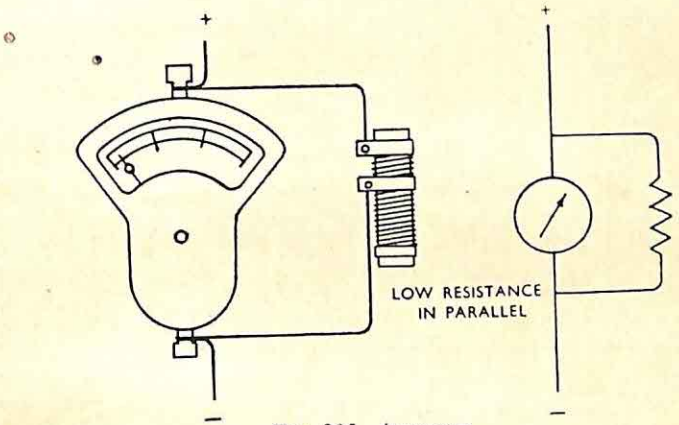


FIG. 205. Ammeter.

instead of 0–15 milli-amps., it is still necessary to limit the maximum current passing through the moving coil to 0·015 amp., and a lower resistance (999 times smaller) must be connected *in parallel*, as shown in Fig. 205, so that *exactly* 0·1 *per cent.* of the current to be measured passes through the moving coil and the remaining 99·9 per cent. is 'by-passed' or 'shunted' through the added resistance. Such a resistance, arranged *in parallel* across the terminals of the instrument, is called a *shunt*.† If the resistance of the moving coil is 999 times greater than that of the shunt, a 15-amp. current will split up so that 0·015 amp. goes through the moving coil and the remaining 14·985 amps. takes the easier path through the shunt, and the instrument gives a full-scale deflection which represents 15 amps.

In this way, a single moving-coil instrument, together with a

set of *high* resistances for connection *in series*, can be used as a *voltmeter* for measuring a wide range of E.M.F.s, and the same instrument, together with a set of *low* resistances for connection *in parallel*, can be used as an *ammeter* for measuring direct currents of any strength.

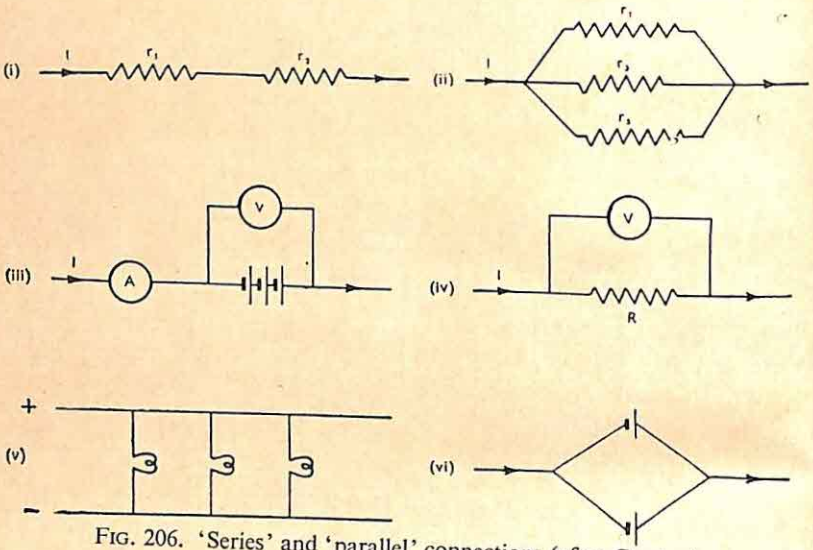


FIG. 206. 'Series' and 'parallel' connections (after Gordon).

CONDUCTORS IN SERIES AND IN PARALLEL

When cells, lamps, resistances, and measuring instruments are connected up in an electrical circuit, they can be arranged in two different ways. If *all the current* passes through each conductor *in turn*, they are said to be '*in series*'. (In Fig. 206 (i), for example, the two resistances r_1 and r_2 are arranged in series.) If they are arranged so that *only part of the current* passes through each of them, they are said to be '*in parallel*'. (In Fig. 206 (ii), for example, the resistances r_1, r_2, and r_3 are arranged in parallel.)

When resistances are arranged *in series*, the total resistance of the series is the *sum* of the individual resistances, i.e. $R = r_1 + r_2$. In the same way, if three dry cells (E.M.F. = 1·5 volts) are con-

nected in series, the total E.M.F. of the combination is 4·5 volts. (Fig. 206 (iii) shows three cells connected in series.)

When resistances are connected *in parallel*, the current is offered a choice of conductors and splits up so that the largest part of the current flows through the smallest resistance, i.e. the current flowing through any one conductor is *inversely proportional to its resistance*. In Fig. 206 (ii), for example, if the equivalent single resistance is R, the current flowing through r_1 is $\frac{R}{r_1}$ of the total current, r_2 carries $\frac{R}{r_2}$ of the total current, and r_3 carries $\frac{R}{r_3}$ of the total current. The relation between the total resistance R and the individual resistances r_1, r_2, and r_3 in parallel is $\frac{1}{R} = \frac{1}{r_1} + \frac{1}{r_2} + \frac{1}{r_3}$. Notice that the sum of a number of resistances in parallel is always *less* than any single resistance, because each additional resistance provides another path through which current can flow.

When several similar *cells* are connected *in parallel* (Fig. 206 (vi)), the total E.M.F. is the same as that of each single cell.

Since an *ammeter* measures the *quantity of electricity* flowing in a circuit, it must be connected *in series* so that *all the current flows through it* (see Fig. 206 (iii)). A *voltmeter*, however, measures only the *potential difference* between two points in the circuit and is therefore connected *in parallel* across these two points (see Fig. 206 (iii)). Perhaps you will find this easier to understand if you think of the air in a motor-car tyre. If we want to measure the *quantity* of air in the tyre we have to let out *all* the air and measure its volume, just as we have to make *all* the current pass through the *ammeter*. If we only want to measure the *pressure* of the air in the tyre, it is not necessary to make all the air pass through the pressure-gauge: only a very little air is used, just as only a very small fraction of the current flows through the *voltmeter* when measuring potential difference (see Fig. 206 (iii) and (iv)).

(See *General Science Workbook*, Book Two, pp. 111–34, for numerical problems and learning exercises on this chapter.)

Chapter XVII

MOTORS, DYNAMOS, AND ELECTRIC POWER SUPPLIES

ELECTRIC MOTORS

If you understand the principle of the moving-coil instrument (pp. 364–6) it is easy to explain how a simple electric motor works. If a current is passed through the moving coil of an ammeter from which the spiral springs have been removed (see Fig. 207), the coil *ABCD* will turn so that its N-pole side faces the S-pole of the permanent magnet, and it will stop in that position as long as the current flows. If the current is then reversed, however, what was the N-pole side of the coil becomes the S-pole, and the coil will turn through 180° so that this S-pole faces the N-pole of the permanent magnet. Hence, *if we can reverse the direction of the current at exactly the right moment, and if we supply current to the coil in such a way that the coil is free to turn, then the coil will go on rotating as long as a current is passed.* Such an arrangement (shown in Fig. 208) would be a simple *electric motor.*

In most practical electric motors *electro-magnets* are used instead of permanent magnets (Fig. 209). These electro-magnets (or *field-magnets*) are made of soft iron and in such a shape as to form a cylindrical gap between the poles. The moving part of the motor is called the *armature*† and it consists of an axle carrying a soft-iron core wound with insulated wire. The ends of each coil of wire in the armature are connected to a *commutator*† (or *current-reverser*). In its simplest form, i.e. with a two-pole armature, this consists of the two halves of a brass ring that revolve with the armature as shown in Fig. 217. Two *brushes* (strips of copper or carbon) press one against each half of the commutator so that as the armature revolves, each brush makes contact with

each end of the coil-winding in turn. In this way the current is reversed as the brushes slip from one half of the commutator to

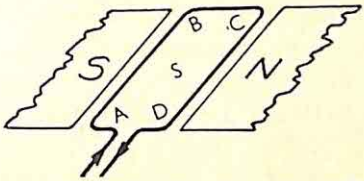


FIG. 207. Principle of moving coil.

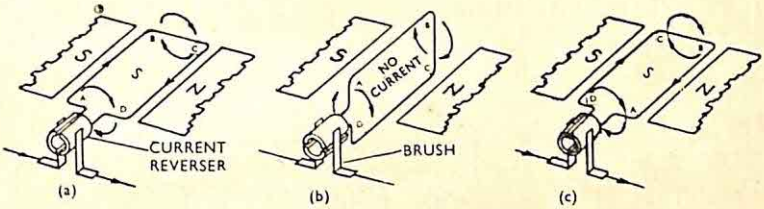


FIG. 208. Principle of D.C. electric motor (after Pearce).

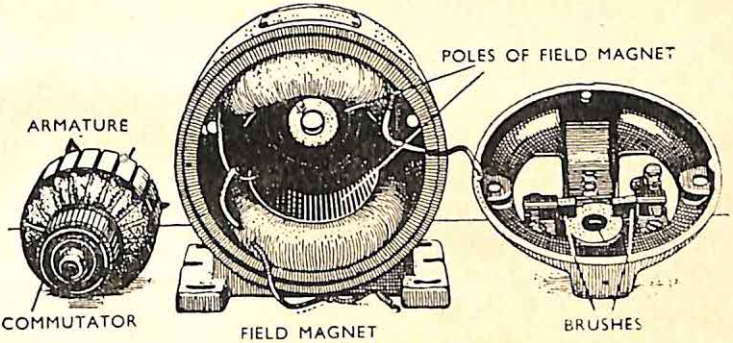


FIG. 209. Practical electric motor.

the other, the current entering through one brush and leaving through the other.

Practical electric motors (see Fig. 209) always have more than one coil in the armature, and the greater the number of coils the smoother the running of the motor.

The current for the field-magnets is taken from the same source as that supplying the armature, but the connections can be made

in two different ways, as shown in Fig. 210, giving either a *series-wound motor* or a *shunt-wound motor*. In a series-motor, the field-magnet coils are in series with the armature-coils, so that the whole current passes through both sets of coils. Series-motors 'pull' hard at slow speeds and are therefore used for electric trains and tram-cars, giving rapid acceleration when starting from rest. Shunt-motors have their field-magnet coils connected in parallel

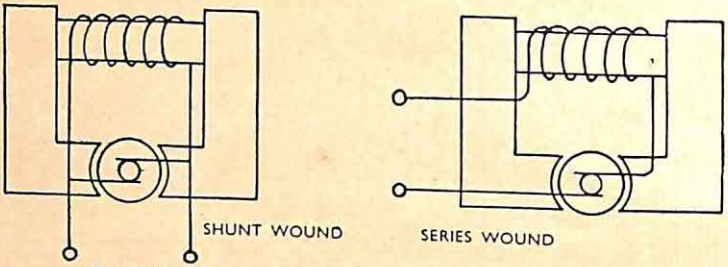


FIG. 210. Electric motor windings (very diagrammatic).

across the armature-coils, so that part only of the current passes through the field-magnet coils and part through the armature. Shunt-motors will maintain nearly the same speed under varying loads and are therefore used for driving machinery, e.g. lathes,† where a fairly steady speed is required under varying loads.

DYNAMOS AND MOTORS

When we supply *electrical energy* to an *electric motor*, the armature revolves and supplies *kinetic energy* that can be made to *do work*. Conversely, if we supply the same motor with *kinetic energy* by turning the armature, it supplies *electricity*. To understand how a dynamo works we must first learn something about *electro-magnetic induction*.

FARADAY'S EXPERIMENTS

(*a*) *Electro-magnetic induction*

In 1831 Faraday, one of the greatest scientists of the nineteenth century, discovered that an electric current was produced when a

bar-magnet was moved towards a coil of wire, and also that when the magnet was drawn away a current was produced, but in the opposite direction (see Fig. 211).

This effect is called *electro-magnetic induction* and an electric current is said to have been *induced* in the coil.

Faraday also found that another coil carrying a current could be used instead of a bar-magnet. When the smaller coil in

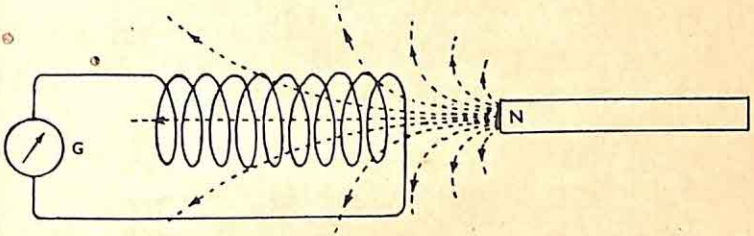


FIG. 211. Electro-magnetic induction.

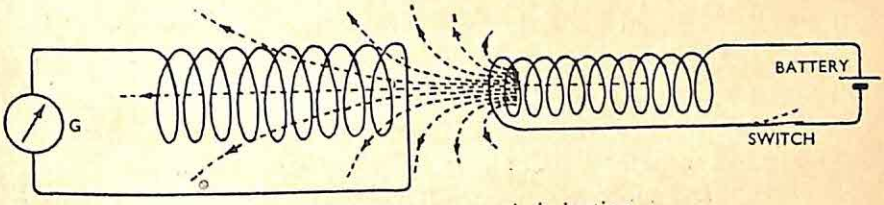


FIG. 212. Electro-magnetic induction.

Fig. 212, carrying the current, is either pushed into or withdrawn from the larger coil, a current is induced in the larger coil.

The same effect can be produced without moving the coils if a switch is included in series with the smaller coil that carries the current. Faraday put a small coil (the *primary coil*) connected to a battery and switch inside a larger coil (the *secondary coil*) connected to a current-detector G. On closing the switch in the primary circuit, the compass-needle in the secondary circuit was deflected, and when the switch was opened again, there was another deflection of the needle, but in the opposite direction. Similarly, when the current was reversed, the deflections of the needle were reversed.

This production of an *induced current* in a circuit when a current flows in a neighbouring circuit is the basis of all dynamos, transformers, and induction coils, and it is used in telephones and in wireless transmitters and receivers. You should remember, therefore, that *the movement of a magnet, or any change in an electric current, always induces currents in neighbouring circuits.*

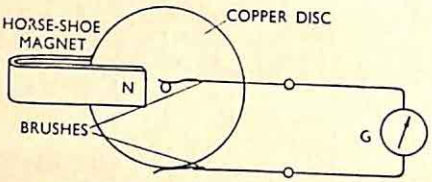


FIG. 213. Magneto-electric effect.

(*b*) *Magneto-electric effect*

During his series of experiments in 1831 Faraday made a copper disc rotate between the poles of a strong horseshoe magnet. The edge of the disc and the axle were in contact with brushes connected to a sensitive current detector (or *galvanometer*—so called after Galvani)† as in Fig. 213.

The galvanometer needle was deflected and indicated a *continuous* induced electric current. This device was the first 'magneto-electric' machine ever made. From it the dynamo was developed.

The same principle is illustrated when a single copper wire is moved through a magnetic field. The wire has a current induced in it *while it is moving* (see Fig. 214).

DYNAMOS (ELECTRIC GENERATORS)

A dynamo is a machine for keeping conductors moving through a magnetic field to produce an electric current.

The principle of the dynamo is explained in Fig. 215. The loop of copper wire ABCD (a simple single coil armature) is rotated between the poles of a circular magnet. When the ends of the loop are connected to two separate rings (*slip-rings* that rotate on the

same axis as the loop OO') a *variable alternating current* flows. This can be tested, using an ammeter. Fig. 216 shows how the current starts from zero with the coil in the position shown in

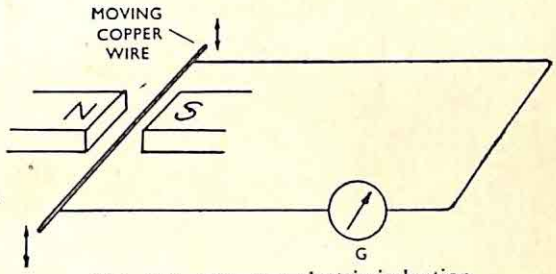


FIG. 214. Magneto-electric induction.

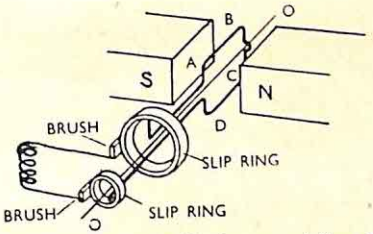


FIG. 215. Principle of A.C. dynamo (after Gordon).

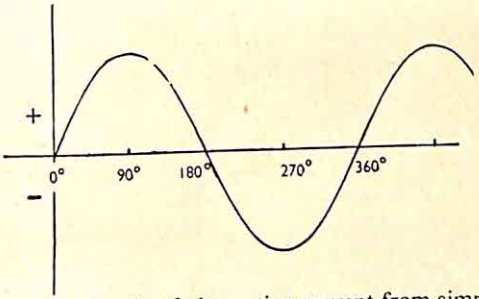


FIG. 216. Graph of alternating current from simple two-pole armature as shown in Fig. 215.

Fig. 215, rises to a maximum (+ve) when the armature has turned through 90°, and falls to zero again after 180°. During the next half-turn the current flows in the opposite direction, increasing to a maximum (−ve) after another 90° and falling to zero again at

the end of the complete turn. The whole cycle of current changes is then repeated during the next revolution.

If a *direct current* (D.C.) is required, a commutator or current-reverser (Fig. 217) must be used to reverse the current every half-revolution of the 2-pole armature. Fig. 218 shows the effect of

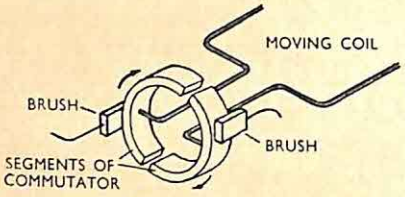


FIG. 217. Principle of commutator or current-reverser (after Gordon).

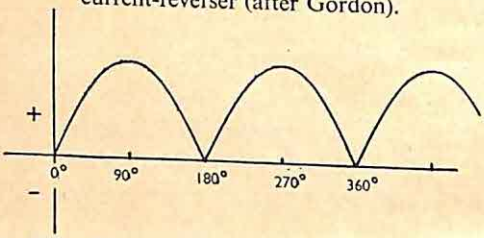


FIG. 218. Graph of direct current from simple two-pole armature fitted with commutator as shown in Fig. 217.

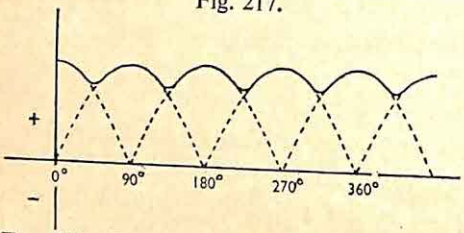


FIG. 219. Graph of direct current from four-pole armature fitted with commutator of four segments.

the commutator in producing a current flowing in one direction only (although not a *steady* current).

Practical dynamos are never made with a single-coil, 2-pole armature, because such an arrangement does not give a steady

current. The larger the number of coils the steadier is the current produced (just as a modern four-cylinder motor-engine gives smoother running than the early single-cylinder engines). Practical dynamos, therefore, have the armature wound with a large number of coils, and have the same number of segments in the commutator. Fig. 219 shows the current produced by a 4-pole armature (i.e. two coils fixed at right-angles to each other). Notice that when the current from one coil is at a maximum, the current from the other coil is at a minimum (as shown by the dotted lines) so that the combined current never falls to zero. The dark line is a graph of the resultant current produced by the double-coil, 4-pole armature, showing that the current is steadier (or 'smoother') than with a single-coil, 2-pole armature.

It is clear that the larger the number of coils, the smoother will be the current.

A *magneto*† is a dynamo with permanent steel magnets producing an alternating current of high voltage. In the small permanent-magnet dynamo used for lighting bicycle-lamps, the coils are usually stationary and the magnets revolve.

THE INDUCTION COIL

For some purposes small currents at a very high voltage are necessary. Thus 'X'-ray tubes and the spark-plugs† of motor-engines require small currents of only a few milli-amperes with a large E.M.F. of thousands of volts. Such currents are most conveniently produced by an *induction coil* (see Fig. 220). The *primary coil*, consisting of a few hundred turns of thick insulated copper wire, is wound on a soft-iron core (usually a bundle of soft-iron wires). The *secondary coil*, consisting of many thousands of turns of very thin insulated copper wire, is wound outside the primary coil. The *primary coil* is connected to a battery in series with a '*make-and-break*' (or current-interrupter) like that used in electric bells (see Book Three, Chap. V). The ends of the *secondary coil* are connected to terminals or two adjustable arms between which sparks can pass.

At each 'make-and-break' (i.e. every time the primary current starts or stops), a current is induced in the secondary coil, but since the 'break' is much more sudden than the 'make', its effect is much greater, and the induction coil gives a rapid succession of

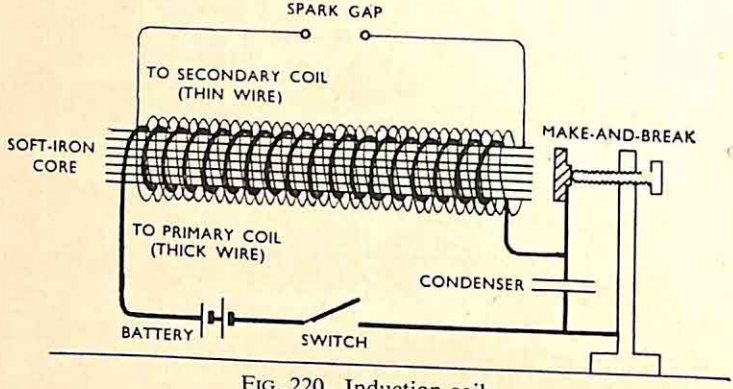


FIG. 220. Induction coil.

rushes of high-voltage direct current instead of the alternating current that would be produced if the 'make' and the 'break' had equal effects.

If the potential difference between the terminals of the secondary coil is big enough, a spark passes between them. In order that a spark may pass between two sharp points one inch apart (in air), a potential difference of about 20,000 volts is necessary. To produce a spark across the normal gap of the spark-plug of a motor-engine requires a potential difference of about 5,000 volts at the pressure inside the cylinder at the top of the compression-stroke. In the open air about 2,000 volts is sufficient.

THE A.C. TRANSFORMER

One of the most important everyday applications of electro-magnetic induction is in *transformers*. Faraday, in 1831, took a soft-iron ring and wound two separate coils of insulated copper wire round it, as shown in Fig. 221. The ends of one coil (the *primary*) were connected to a battery through a switch, while the ends of

the other coil (the *secondary*) were connected to a current-detector. When the switch in the primary circuit was closed the compass-needle in the secondary circuit was deflected for a moment, but returned to its original zero position until the switch was opened

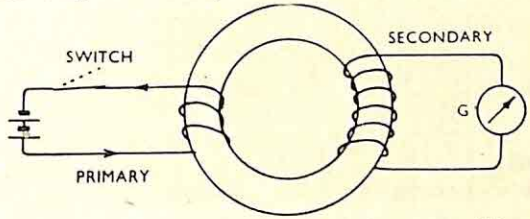


FIG. 221. Faraday's transformer (diagrammatic).

again. The needle was then deflected in the opposite direction for a moment.

It is important to notice that there is no induced current in the secondary coil as long as the primary current is steady, *but only*

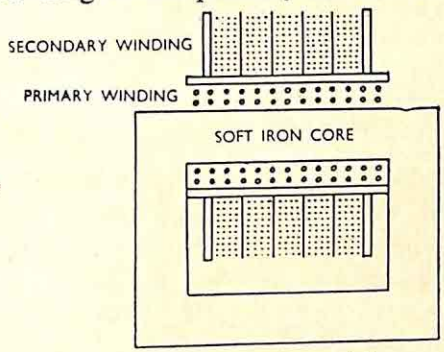


FIG. 222. Section through transformer (diagrammatic).

when it changes. An *alternating current*, however, is rapidly changing in strength and direction (see Fig. 216), hence if an alternating current is passed through the primary coil, an alternating current of the same frequency will be induced in the secondary coil. This is the principle of the *transformer* that is used in the distribution of electrical power. Practical transformers consist of a primary and a secondary coil of insulated wire wound on a soft-iron core as shown in Fig. 222, and by varying the number of turns in the

primary and secondary windings, an alternating current can be '*stepped up*' or '*stepped down*' to any required voltage.

If the primary and secondary coils have the same number of turns, the induced E.M.F. in the secondary coil will be the same as the E.M.F. applied to the primary coil. If the secondary coil has twice as many turns as the primary, then the E.M.F. induced in the secondary will be twice that in the primary coil. In general:

$$\frac{\text{Secondary voltage}}{\text{Primary voltage}} = \frac{\text{number of turns in secondary coil}}{\text{number of turns in primary coil}}.$$

Hence, by varying the number of turns in the primary and secondary windings of a transformer, we can change the voltage of an alternating current supply to suit any requirements. A 'step-up' transformer has fewer turns in the primary than in the secondary, and it 'steps up' the supply from a lower to a higher voltage. A 'step-down' transformer has more turns in its primary coil than in its secondary, and it 'steps down' a high-voltage alternating current to a lower voltage.

We cannot, however, get more *electrical energy* out of the secondary coil than we put into the primary coil; i.e. the *output* (in watts) is equal to the *input* (in watts), if we neglect small heat losses, and what we gain in *voltage* we lose in *current* (and vice versa). For example, suppose a transformer has 1,000 turns in the primary to every 100 turns in the secondary (i.e. it is a 10-to-1 step-down transformer), and the input to the primary is a current of 1 amp. at 100 volts (= 100 watts). Then the output from the secondary will be a current of 10 amp. at 10 volts (= 100 watts), and we have transformed a small current at a high voltage into a large current at a low voltage.

TRANSMISSION OF ELECTRICAL POWER

In most of our elementary lessons on electricity we have dealt with the effects of direct currents (i.e. currents flowing steadily in one direction) but all modern power supplies are *alternating current* (i.e. a current flowing backwards and forwards many times a

second—50 cycles per second in most public lighting supplies). Although A.C. is as good as D.C. for driving motors, heating, and lighting, it is useless for electrolysis or for charging accumulators, and it gives much more dangerous shocks than D.C. at the same voltage. The fundamental reason why most power-stations supply A.C. is because transformers can be used with A.C. but not with D.C. An alternating current can be transformed to any required voltage, either up or down, with very little loss of energy. With D.C. it is a much more difficult and wasteful operation to change the voltage.

The necessity for changing the voltage of power supplies arises in connection with the *long-distance transmission* of electric currents. We have seen (p. 347) that some heat is produced whenever a current flows through a conductor. The rate at which energy is wasted in useless heating of electric transmission wires depends only on two factors: (*a*) the *resistance* of the wires, and (*b*) the *current* flowing through them (the heat produced being proportional to the *square* of the current). The resistance of the wires is kept low by making them of copper or aluminium. The current is kept low by raising the voltage, and since the quantity of electrical power (in watts) transmitted is the product of the current (in amperes) and the voltage (in volts), the same amount of electrical *energy* is carried by a small current at a high voltage as by a large current at a low voltage, and the loss of energy will be much less with the smaller current.

To take a practical example: Suppose it is required to supply 230 Kilowatts from a power-station to a factory 3 miles away. Let us compare the relative cost and efficiency of transmitting this energy (*a*) at 230 volts and (*b*) at 6,000 volts. (See Table on p. 386.)

Notice that by increasing the voltage 26 times the transmission losses are reduced 26 times. Besides this, the cost of copper is about 26 times less when the higher voltage is used.

Transformers are used for stepping up the voltage of power supplies for long-distance transmission. Such high voltages, however, are very dangerous and have to be stepped down by other

	(*a*)	(*b*)
Distance for transmission	3 miles	3 miles
Electrical energy to be supplied	230,000 watts	230,000 watts
Voltage of supply at factory end	230 volts	6,000 volts
Current required (watts ÷ volts)	1,000 amps.	38·3 amps.
Cross-section of wire required to carry this current without overheating	1 sq. in.	0·04 sq. in.
Approximate weight of wire required	29 tons	1·1 tons
Approximate cost of this copper wire	£7,910	£300
Resistance of wire	0·125 ohms	3·265 ohms
'Voltage Drop' (Ohm's Law: E = C×R)	0·125×1,000 = 125 volts	38·3×3·265 = 125 volts
Voltage needed at power-station end	230+125 = 355 volts	6,000+125 = 6,125 volts
Energy wasted in trasmission by useless heating of wires. (Voltage Drop×Current)	125×1,000 = 125 Kw.	125×38·3 = 4·8 Kw.
Relative Efficiency of Transmission	45·6%	97·9%

transformers to give a lower voltage that will be safe for household use. To take an actual example: The output of a power-station in a large town is at 6,600 volts. For transmission to another town 20 miles away, this is stepped up to 66,000 volts. On arrival at the second town it is stepped down to 11,000 volts for transmission to the 'sub-stations' in different parts of the town, from whence it is distributed to houses at 230 volts, this being the highest voltage that is at all safe for household use.

To change a direct current from one voltage to another involves feeding an electric motor with the first current and making this motor drive a dynamo producing current at the required voltage, a much more complicated, expensive, and wasteful process than changing the voltage of an alternating current with a transformer.

Another advantage of transformers is that when the primary winding is connected to a supply of A.C., only a small current passes when the secondary circuit is 'open' and no current is being taken from it. The primary current is said to be 'choked'. When

the secondary circuit is 'closed' and current is taken from it, the primary winding is no longer 'choked' and a larger current flows through it. For example, a *bell-transformer*, used instead of dry cells for ringing electric bells, steps down 230 volts A.C. to 6 or 8 volts, and can be left permanently connected to the electric light mains because the primary winding takes a very small current indeed until the secondary circuit is closed by pressing a bell-push. Then, as soon as a current flows in the secondary winding, the primary circuit takes its full current from the mains.

CONVERSION OF A.C. TO D.C.

There are several methods of converting A.C. to D.C. One method is to feed the A.C. to an A.C. motor driving a D.C. dynamo. This arrangement is called a *rotary** *converter* and it gives a *smooth* D.C. output.

Another method is to use an *electric valve*, which allows current to flow in one direction but not in the other.

Such a *valve-rectifier* converts (or *rectifies*†) an A.C. into pulses of D.C. This pulsating D.C. can be 'smoothed' by passing it through a '*choke*' or 'choking-coil' (which is simply a transformer with only one coil).

Small D.C.s at a low voltage are very often required in everyday life, e.g. for charging accumulators and for wireless sets, and these are most easily obtained by using a *copper oxide rectifier*, which consists of discs of copper covered with a layer of copper oxide on one side only. Currents flow through such a copper-copper oxide contact much more readily in one direction than in the other, so that when an A.C. is passed, if the positive half of the wave flows across the copper-copper oxide boundary, the negative half is stopped. In this way an A.C. is *rectified* to produce a pulsating D.C. (which can be 'smoothed' by passing through a 'choke').

Chapter XVIII

TELEGRAPHS AND TELEPHONES

We have seen that one of the most important everyday uses of electric currents is to convey electrical *energy* to a distant point where it can be converted into other forms of energy. Almost equally important is the use of electric currents to convey *messages* over long distances. Before the invention of telegraphs and telephones, news and messages had to be written and sent by land or sea, so that they travelled with no greater speed than Man himself (except, perhaps, when carrier-pigeons were used). Using telegraphs and telephones, messages arrive at a distant 'receiver' at almost the same instant as they are sent by the 'transmitter'.

The *telegraph* was the earliest instrument to use an electric current for carrying messages. One of the simplest types uses an electro-magnet at the receiving end, worked by currents sent along a wire by the *tapping-key* at the transmitting end. When the tapping-key is pressed, the circuit is closed and a current flows along the wire and through the electro-magnet at the receiving end, attracting a piece of soft iron that either makes a sharp sound or presses an inked wheel against a moving strip of paper. By using the Morse† Code,* where each letter of the alphabet and every number is represented by a combination of 'dots' and 'dashes', a message can be transmitted as a series of impulses of current, a short impulse being a 'dot' and a longer impulse a 'dash'.

In modern *teleprinters*† the operator taps out the message on a transmitting instrument that looks like a typewriter and which sends electrical impulses along the wires to the receiving end where a similar machine types out the message on a strip of paper. In telegraph-offices this strip of paper is cut up and gummed on a telegram form.

Nowadays, the telegraph is used mainly for long-distance transmission, and the *telephone* is a more familiar instrument in daily life. In the telephone *transmitter*, sound-waves (i.e. waves of compression and rarefaction, or variations of air-pressure) produce

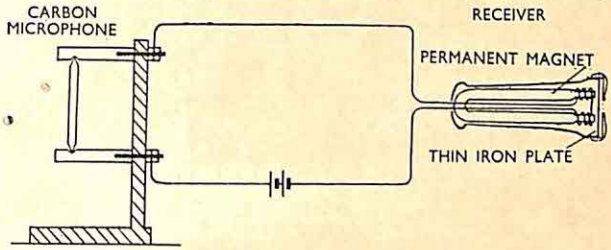


FIG. 223. Simple carbon microphone and telephone receiver.

variations of electric current that travel to the *receiver* where these variations of current are converted once more into variations of air-pressure (or sound-waves).

To understand how the telephone transmitter works, you should first examine a simple *carbon microphone*,† consisting of a hard carbon rod, sharpened at both ends and held loosely between two carbon plates as shown in Fig. 223. The electrical resistance at these carbon contacts depends on the force pressing them together, and anything that changes the pressure on these contacts also changes the current flowing in the circuit. Sound-waves will do this, so that the microphone will convert variations of air-pressure into variations of electric current.

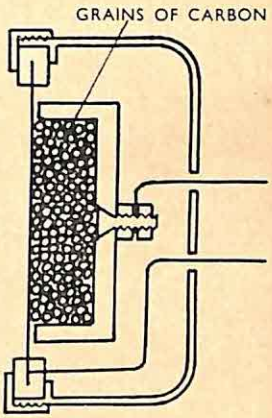


FIG. 224. Practical telephone transmitter.

If a watch is placed near such a carbon microphone, connected in series with a dry battery and a telephone receiver, the ticking of the watch can be heard in the receiver.

In practical telephone transmitters like the one shown in Fig. 224,

a large number of grains of carbon are used instead of the two contacts of the simple microphone, and the thin plate (or *diaphragm*) vibrates backwards and forwards 'in step' with any sound-waves falling on it. When the diaphragm is pushed inwards by a wave of compression, the grains of carbon are pressed

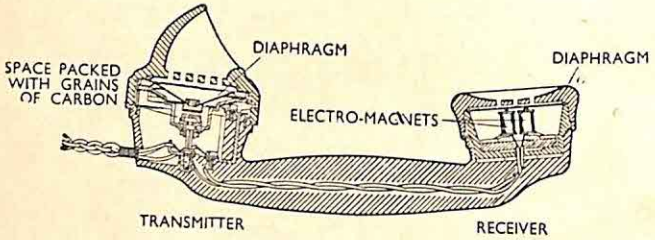


FIG. 225. Telephone 'handset'.

more tightly together and more current flows, and when the diaphragm moves outwards during a wave of rarefaction, the current is weaker. Hence the variations in air-pressure caused by the sound-waves are converted into variations of electric current by the transmitter.

In the telephone *receiver* this varying current passes round coils of wire surrounding the poles of a permanent magnet that attracts a thin iron plate (see Fig. 223). Variations in this current produce corresponding variations in the strength of the magnet, so that when a wave of compression strikes the diaphragm of the *transmitter* and a stronger current flows round the coils of the *receiver*, the iron plate is pulled inwards, and when a wave of rarefaction at the transmitter produces a weaker current, the iron plate in the receiver flies back, thus vibrating 'in step' with the diaphragm of the transmitter. The vibrations of the diaphragm in the receiver set up sound-waves corresponding to those reaching the transmitter. In most modern telephones the transmitter and receiver are both fitted into a bakelite *handset*, so that when the receiver is held to the ear the transmitter is opposite the mouth. Fig. 225 shows a common type of combined instrument.

Fig. 226 shows a practical telephone circuit. Only low voltages can be used in the microphone-circuit (not more than 3 volts), otherwise the carbon contacts burn and crackling noises are pro-

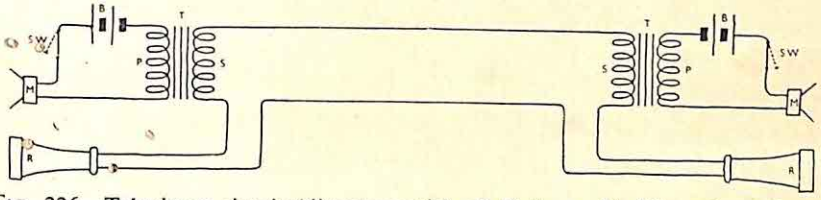


FIG. 226. Telephone circuit (diagrammatic). B. Battery. M. Microphone (transmitter). R. Receiver. T. Step-up transformer. P. Primary winding. S. Secondary winding. SW. Switch.

duced, hence the low-voltage current from the transmitter is stepped up for transmission through the long wires leading to the distant receiver (just as in power-transmission) to reduce losses due to line-resistance.

Chapter XIX

THE NATURE OF ELECTRICITY AND THE STRUCTURE OF MATTER

In earlier lessons we learnt that matter is made up of very small particles called *molecules*: a molecule being the smallest portion of a substance that possesses the characteristic properties of that substance. Molecules themselves are built up of chemical *atoms*: an atom being the smallest portion of an element that can take part in a chemical change. Up to the end of the nineteenth century the *atom* was thought to be the smallest particle of matter, but at the beginning of the twentieth century it was found that atoms are not solid and indivisible; they are built up of still smaller particles.

In 1910 Rutherford, the great British physicist, showed that an atom consists of an *atomic nucleus*, carrying a positive charge of electricity, surrounded by particles of negative electricity that he named *electrons*. This arrangement reminds us of the solar system with its planets revolving round the central Sun. The number of electrons revolving round an atomic nucleus depends on the number of charges of positive electricity carried by the nucleus, so that in every atom the positive and negative charges just balance, making the atom *neutral*, or uncharged. Thus, the nucleus of a hydrogen atom has only one unit of positive electrical charge and has only one electron revolving around it: uranium (the heaviest natural element) has 92 units of positive charge and its nucleus has 92 electrons revolving round it. Between the lightest and the heaviest natural atoms (hydrogen and uranium) are all the various combinations of positive and negative charges that exist in Nature and which give rise to all the different kinds of atoms. The differences between the chemical properties of different elements are due to the number and arrangement of the

electrons in the different kinds of atoms, in fact, the 92 elements that are found in Nature can be arranged in a *periodic table* according to the number of electrons in their atoms, e.g. hydrogen has 1 electron revolving round its atomic nucleus, helium has 2, lithium 3, beryllium 4, boron 5, carbon 6, nitrogen 7, and so on, up to uranium with 92 electrons. Removal or addition of outer electrons causes the atom to become a charged *ion* (see p. 353).

Later work showed that the *atomic nucleus*, too, has a structure (except in the case of hydrogen, where the nucleus consists of a single particle of positive electricity). In all the other 91 natural elements, the atomic nucleus contains two kinds of particles—*protons* and *neutrons*. (Rutherford discovered and named the proton; his former pupil, Chadwick, discovered and named the neutron.) Protons and neutrons are about equal in mass or weight but they are nearly 2,000 times as heavy as electrons. Electrons can be moved much more easily than protons or neutrons. Protons and neutrons have quite different electrical properties: *the proton is positively charged*, and the number of electrical charges carried by the atomic nucleus is equal to the number of protons it contains. (This number is called the *atomic number* of the element; see the table on p. 398.) Thus, the hydrogen atom with one positive charge has one proton and its atomic number is 1; uranium, with 92 positive charges, has 92 protons in its nucleus, and its atomic number is 92. *The neutron*, however, *is electrically neutral* and carries no charge at all. Because of this absence of electric charge, the neutron can pass readily through matter. But its mass or weight is very nearly the same as that of the proton, and the weight of an atom is the sum of the protons and neutrons in its nucleus (its electrons are so light that their mass or weight can be ignored when considering the weight of an atom). Oxygen has 8 protons and 8 neutrons in its nucleus and its atomic weight is 16.

ATOMIC ENERGY

In the nucleus of the atom, these protons and neutrons are held

together by the most powerful forces known to scientists; more than a million times as strong as the chemical forces that bind together the atoms of hydrogen and oxygen in the water molecule (and these chemical forces are much stronger than the *cohesion* between the particles in a piece of hard steel). The strength of these forces holding protons and neutrons together in the atomic nucleus can be calculated. It is found that the combined mass of all the protons and neutrons inside any atomic nucleus is about 1 per cent. less than the sum of the masses of the individual protons and neutrons because *some of the mass of the individual particles has been converted into energy*.[1] This energy difference is the strength of the forces holding the protons and neutrons together in the atomic nucleus. It is clear that the same amount of energy would have to be supplied in order to tear the nucleus apart into its protons and neutrons, hence such 'atom-splitting' would not serve as a source of energy. With one exception, no atomic nucleus has been completely split into separate protons and neutrons, although many atomic nuclei have been split into two parts, one very small indeed and the other containing all the rest of the nucleus. In 1936, however, scientists succeeded in splitting a uranium nucleus into two nearly equal parts weighing together about 0·1 per cent. less than the original material and thus releasing an equivalent amount of energy. This type of 'atom-splitting' is called *nuclear fission*† and is the basic principle of the atomic bomb. This *fission* of a uranium atom is brought about when a neutron hits the nucleus. During this fission, neutrons are

[1] Now that you know more about the structure of matter you will understand why on pp. 2–3, in stating the *Law of Conservation of Weight*, we were careful to say that there is no loss of mass or weight *in any chemical change*. The smallest particle that takes part in a *chemical change* is the atom, but you will now realize that when we come to consider *nuclear changes* inside the atom the two great scientific principles called *conservation of energy* and *conservation of mass* overlap and that *mass can be converted into an equivalent amount of energy* (in accordance with Professor Einstein's law of the equivalence of mass and energy). When protons and neutrons come together to form an atomic nucleus, that fraction of the mass of the individual particles which disappears is not 'destroyed' but is converted into an equivalent amount of energy. For example, one ounce of mass is equivalent to the energy produced by the biggest power station in the world during a whole month.

set free and these, striking other uranium nuclei, produce further fission and, as the liberated neutrons multiply, more and more atomic energy is set free by the *chain reaction* until all the uranium present has been split into lighter portions. Such an uncontrolled *nuclear chain reaction* produces the violent explosion of the atomic bomb.

During *chemical changes*, only the loosely-bound electrons in the outer part of the atom are rearranged, but some of these chemical changes, e.g. the burning of wood or coal or oil, supply great quantities of energy that we use in everyday life. In such chemical changes, however, countless millions of atoms react together in a short time and the energy set free per atom is very small indeed. During *nuclear changes*, it is not the loosely-bound electrons that are changing position, but the protons and neutrons of the nucleus. We have seen that these are bound together by enormous forces, hence, when they change their positions, enormous amounts of energy are set free. The energy liberated by the nuclear fission of a single atom is about 100 million times greater than that set free by a single atom undergoing chemical change. Practically all the energy that is used on the Earth for power comes, in the first place, from atomic nuclear changes taking place in the Sun. Fission chain reactions are now being used to produce energy for industrial use. Fortunately, the chain reaction can be kept under control and this enables Man to use atomic energy for producing useful power instead of destructive explosions.

ISOTOPES

It was stated on p. 2 of this book that some of the points in *Dalton's Atomic Theory* have had to be modified in the light of recent discoveries, e.g. it is now known that all the atoms of the same element are *not* exactly alike. It has been found that almost every chemical element has atoms of differing atomic weight. Such atoms of *the same element* but of *different weight* are called *isotopes*, and the natural element consists of a mixture of these isotopes. For example, carbon has two isotopes, nitrogen has two,

oxygen has three, uranium has four. Different isotopes of the same element have *different numbers of neutrons* in their atomic nuclei (although *the number of protons in the atoms of any one element is always the same*). In the case of uranium, each atom (of any isotope) has exactly the same number of protons—92. But there is one kind of uranium with an atomic weight of 234, another of atomic weight 235, another 238, and another 239. Since each has exactly 92 protons in each atom, the difference in atomic weight can only be due to differences in the number of neutrons. Uranium-238 has 146 neutrons in its nucleus; uranium-235 has only 143. U-235 is the rare, explosive isotope, used in atomic bombs; U-238 is not explosive and is much more plentiful. The separation of these uranium isotopes, found mixed together in Nature, is very difficult and can only be carried out in large industrial plants.* Similarly, most of the hydrogen in the world has an atomic weight of 1 because its nucleus consists of a single proton. But there is a small proportion of natural hydrogen atoms that also have a neutron in their nucleus, doubling their atomic weight. Hence the atomic weight of natural hydrogen shown on p. 398 (1·008) is the average weight of a very large proportion of H-1 ('light' hydrogen) and a very small proportion of H-2 ('heavy' hydrogen).

ELECTRIC CHARGES AND CURRENTS

Electrically charged objects have either an excess of or a shortage of electrons, e.g. ebonite, when rubbed with fur, *gains* electrons that have been removed from the fur, and thus the ebonite becomes negatively charged; glass, on the other hand, when rubbed with silk, *loses* electrons to the silk, and thus the glass becomes positively charged.

Insulators, i.e. substances that do not allow electricity to flow through them, have very stable arrangements of electrons in their atoms, so that the electrons are not readily disturbed.

Conductors, i.e. substances that allow electricity to flow

through them, have electrons that are readily disturbed and separated from their atoms.

The action of a battery is simply to pile up electrons at the negative pole. Consider a single electron at the negative pole. When this pole is joined by a conductor to the positive pole, this electron rushes into the nearest atom of the conductor, drives out another electron, and takes its place. This displaced electron repeats the process with the next atom of the conductor, and so on with succeeding atoms along the conductor. This goes on at the rate of several million exchanges per second, and this movement of electrons is the *electric current*. Different atoms allow this displacement of electrons with different degrees of opposition, and this opposition to the displacement of electrons is called the *resistance* of the conductor.

A *dynamo* is simply a device for making electrons move, and, in general, electrons may be set in motion (*a*) by *heat*, (*b*) by *chemical action*, and (*c*) by *electrical charges*.

TABLE OF THE ELEMENTS

with atomic numbers (in brackets before the name), symbols, and atomic weights.

(89)	Actinium	Ac	227
(13)	**Aluminium**	Al	26·97
(51)	**Antimony**	Sb	121·76
(18)	**Argon**	A	39·944
(33)	**Arsenic**	As	74·91
(56)	**Barium**	Ba	137·36
(4)	Beryllium	Be	9·02
(83)	**Bismuth**	Bi	209·0
(5)	**Boron**	B	10·82
(35)	**Bromine**	Br	79·916
(48)	Cadmium	Cd	112·41
(55)	Caesium	Cs	132·91
(20)	**Calcium**	Ca	40·08
(6)	**Carbon**	C	12·01
(58)	Cerium	Ce	140·13
(17)	**Chlorine**	Cl	35·457
(24)	**Chromium**	Cr	52·01
(27)	**Cobalt**	Co	58·94
(29)	**Copper**	Cu	63·57
(66)	Dysprosium	Dy	162·46
(68)	Erbium	Er	167·2
(63)	Europium	Eu	152·0
(9)	**Fluorine**	F	19·00
(64)	Gadolinium	Gd	156·9
(31)	Gallium	Ga	69·72
(32)	Germanium	Ge	72·60
(79)	**Gold**	Au	197·2
(72)	Hafnium	Hf	178·6
(2)	**Helium**	He	4·003
(67)	Holmium	Ho	164·94
(1)	**Hydrogen**	H	1·008
(61)	Illinium	Il	146
(49)	Indium	In	114·76
(53)	**Iodine**	I	126·92
(77)	Iridium	Ir	193·1
(26)	**Iron**	Fe	55·85
(36)	**Krypton**	Kr	83·7
(57)	Lanthanum	La	138·92
(82)	**Lead**	Pb	207·21
(3)	**Lithium**	Li	6·94
(71)	Lutecium	Lu	174·99
(12)	**Magnesium**	Mg	24·32
(25)	**Manganese**	Mn	54·93
(80)	**Mercury**	Hg	200·61
(42)	Molybdenum	Mo	95·95
(60)	Neodymium	Nd	144·27
(10)	**Neon**	Ne	20·183
(28)	**Nickel**	Ni	58·69
(41)	Niobium[1]	N	92·91
(7)	**Nitrogen**	N	14·008
(76)	Osmium	Os	190·2
(8)	**Oxygen**	O	16·000
(46)	Palladium	Pd	106·7
(15)	**Phosphorus**	P	30·98
(78)	Platinum	Pt	195·23
(84)	Polonium	Po	210
(19)	**Potassium**	K	39·096
(59)	Praseodymium	Pr	140·92
(91)	Protactinium	Pa	231
(88)	Radium	Ra	226·05
(86)	Radon[2]	R	220·0
(75)	Rhenium	Re	186·31
(45)	Rhodium	Rh	102·91
(37)	Rubidium	Rb	85·48
(44)	Ruthenium	Ru	101·7
(62)	Samarium	Sa	150·43
(21)	Scandium	Sc	45·10
(34)	Selenium	Se	78·96
(14)	**Silicon**	Si	28·06
(47)	**Silver**	Ag	107·88
(11)	**Sodium**	Na	22·997
(38)	**Strontium**	Sr	87·63
(16)	**Sulphur**	S	32·06
(73)	Tantalum	Ta	180·88
(43)	Technetium[3]	Tc	100
(52)	Tellurium	Te	127·61
(65)	Terbium	Tb	159·2
(81)	Thallium	Tl	204·39
(90)	Thorium	Th	232·12
(69)	Thulium	Tm	169·4
(50)	**Tin**	Sn	118·70
(22)	Titanium	Ti	47·90
(74)	Tungsten	W	183·92
(92)	Uranium	U	238·07
(23)	Vanadium	V	50·95
(54)	**Xenon**	Xe	131·3
(70)	Ytterbium	Yb	173·04
(39)	Yttrium	Yt	88·92
(30)	**Zinc**	Zn	65·38
(40)	Zirconium	Zr	91·22

[1] or Columbium. [2] or Niton. [3] or Masurium.

GLOSSARY

(Simple definitions of general-purpose words in the text that are not included in the standard vocabulary on which the Course is based and which were not used in Books I, II, and III.)

annual, (here) plant that completes its life-cycle, from germination to seed production, within one single year or growing season, 216
aqueous, watery, dissolved in water, 43

benzine, motor-spirit, petrol, solvent for 'dry-cleaning', 67 (not to be confused with *benzene*)
billion, (here) a million million, 10^{12} (British usage) and *not* a thousand million, 10^9 (American usage), 326
blast (*n.*), strong current of air used in smelting, 44; also sudden wave of compressed gas produced by *blasting* with explosives, 29
blunder (*v.*), move or act blindly, find out by chance, also *n.*, 213
brewer, man who *brews* or makes beer,† 84

cancel out, said of, e.g., balancing quantities on both sides of an equation that can be struck out, 13
carbon paper, thin paper coated with carbon for taking copies of letters, etc., 57
cartridge, charge of explosive for guns or *blasting* packed in cylinder of paper, metal, etc., 29
choke (*v.*), stop the supply of air for breathing, etc., 34; also *n.*, choking† coil used in electrical circuits, 386–7.
code, set of letters or figures or words given special meanings for shortness or secrecy, 388
component, constituent part, 261
confine, keep inside, keep within limits, 29
consume, use up, destroy, 327
counter (*adv.*), against, opposite to; hence *counter-balance* (*v.*), and *counterpoise* (*n.*), weight used to balance another weight, also *v.*, 306
crater, hole left in the top of a mountain after *volcanic* explosion, 330

dentist, person skilled in the care of teeth, 28
derive, obtain from, 93
diagonal (*n.*), straight line joining opposite corners of a rectangle, 261
disinfectant, substance that kills germs and makes free from infection, 35
dislocate, put (a bone) out of joint, 184

dispersal, scattering outwards, 233
dough, paste of flour and water for making bread, 84
duration, time that a thing lasts or *endures*, 163
dwarf, animal or plant much below normal size, 211

eliminate, get rid of, *expel*, 168
ellipse, regular oval shape, 318
evacuate, empty out, clear out, discharge (faeces, etc.), 116
expel, throw out from somewhere or something, 136

fatigue, tiredness, 193
feature, notable or characteristic part of thing, 47
flannel, smooth-surfaced, loosely-woven woollen cloth, 333
forge (*v.*), shape by heating and hammering, 41
formulate, express systematically, e.g. as a formula or theory, 214
fragment, part broken off, 59

glide, change position by continuous smooth movement, 182

hearth, floor of fireplace or furnace, 44

identity, absolute sameness, *adj.*, *identical*, 53
ignite, set on fire, take fire, *n.*, *ignition*, 53
indestructible, cannot be destroyed, also *n. indestructibility*, 2
inherit, receive (characteristic, property, etc.) from parents, also *n. inheritance*, 254
initial (*adj.*), at the beginning, first, earliest, 218
insensible, unconscious, also *n. insensibility*, unconsciousness, 28
intoxicate, make drunk, 81
investigate, examine, inquire into, 146

laden, loaded, 198
latitude, angular distance from Equator, to north or south, 319

magnitude, size, of a line, area, volume, angle, force, or any physical quantity, 298
main (here), principal pipe (or wire) for the supply of water (or electricity), 61
mature, fully developed, ripe, adult, 102
mosaic, picture or pattern made by laying and fixing side by side small differently coloured pieces of glass or stone, 179

navigation, science of directing ships at sea, 324

observatory, building from which the sky is studied, 317
obstacle, something that hinders movement or progress, 269

pant (*v.*), breathe rapidly and noisily through open mouth, 144
parchment, thin sheep-skin prepared for writing on; kind of toughened paper similar to this, 79
passive, acted upon, not active, 231
peak, pointed top, e.g. of mountain, of graph, etc., 330
pellet, small ball (or *pill*) of material, 156
perceive, become conscious of through one of the senses, 178
perpetual, going on for ever, 269
phase, one particular stage in a cycle of changes, 330
plant (here), large factory where many different industrial processes are carried out, 396
posture, position of the body, 197
precise, exact, strictly worded, 287
predict, foretell, 269
preliminary, leading up to, preparing for, 43
prolong (*v.*), make longer, cause to continue, 151
prominent, projecting, also *n. prominence*, 328
propel, push forward, also *n. propulsion*, propelling, 271 and *propellent*, something that propels, e.g. explosive supplying force behind bullet, etc., 29

qualitative, concerned with describing kinds or properties of things, 34
quantitative, concerned with measuring quantities of things, 2

reciprocal, in return for, equivalent (see also Fig. 1, page 6, and explanation in text), 2
refine, purify, 31
reject, not accept, put aside without using, 168
roast, expose to strong dry heat in oven or furnace, 31
rotary, rotating, acting by rotation, 387

scum, impurities that have risen to surface of liquid, 90
secrete, (here) produce something by *secretion*, 108
secretion, process of passing out product of gland-cells; this product is also called *a secretion* (e.g. bile from liver), 108
shatter, break suddenly into pieces, 29
sludge, soft thick mud, 66
stable (*adj.*), not easily moved or changed or destroyed, steady, 32
sterile, free from living germs, *v. sterilize*, make free from living germs, 164
still (*n.*), distilling apparatus consisting of boiler and condenser, 69

strain (mech.) (*n.* and *v.*), change in shape and size of a body, caused by forces acting on it, 223
surplus, amount left over when needs have been satisfied, 172
symbol (chem.), a letter used in chemical formulae and equations to represent one atom of a chemical element, 9
synthetic (substance), substance synthesized by chemists, not made by Nature, 26

tangential, in the position or direction of a tangent, 219
textiles, woven materials, cloth, 39
thermostat, instrument for maintaining a constant temperature, 146
tile, thin flat or curved piece of baked clay, 80
twinkle (*v.*), sparkle, shine with an unsteady light, 325

unstable, opposite to *stable*, 32
Universe, The, all the stars and planets, everything everywhere, 295

volcanic, of a *volcano*, mountain or hill with funnel or *crater* through which steam, ashes, melted rock, etc., are *expelled*, 29

wick, twisted or woven cotton that raises oil (or melted wax) by capillary action in oil-lamp (or candle), 70
wriggle (*v.*), move with twisting, jerking turns, 135

INDEX

Where there are several different page-references to the same subject, the principal reference is printed in heavier type. The dagger-sign (†) in front of a word shows that this is a technical word that has not been used before in Books I, II, and III, and which is not explained in the Glossary. These new technical words are also marked (†) where they first appear in the text, and where their meaning is explained.

absolute alcohol, 82.
absorption, 109.
A.C., **378–9**, 383.
†acceleration, 262.
— due to gravity, 267.
accumulator, lead, 344–5.
†acetates, 86.
†acetic acid, 55, **86–87.**
†acetone, 55.
acetylene, 66–67.
acidity, 353–4.
acid sodium sulphate, 27.
†adrenal body, 211.
†adrenalin, 211.
†aerobic respiration, 155.
†after-damp, 65.
air-sacs, in lungs of mammals, 140–1.
— in birds, 139–40.
albumen, 243, **247.**
alcohol, absolute, 82.
— ethyl, 81.
— methyl, 55, **82.**
— wood, 55, **82.**
alcoholic fermentation, 81–86.
alkali manufacture, 38–40.
alkalinity, 353–4.
alkalis, 38–40.
†allotrope, 31.
allotropic carbon, 55.
— phosphorus, 52–53.
— sulphur, 31–32.
†alloy, 42–43.
aluminium, 41.
aluminium manufacture, 356.
alternating current, **378–9**, 383.
American sulphur, 31.
†amino-acetic acid, 93.
†amino-acids, **93–94**, 114–15, 119, 124.
†amino-group, 93.
†ammeter, 357, **369–72.**
ammonia, **21–27**, 61–62.
— soda process, 38–39.
— synthetic, 26–27.
†ammonium bicarbonate, 38–39.
— hydroxide, 23–24.
— salts, 23.
Amoeba, excretion in, 168.
— movement in, 180.
— nutrition in, 131–2.
— reproduction of, 238.
— respiration in, 134.
— sensitivity of, 194.
†ampere, 357.
†amylase, **107**, 111, 113, 123.
†anabolic changes, 101.
†anaerobic respiration, 155.
†anhydrous sodium carbonate, 39.
animal calorimeter, 148.
— charcoal, 55, **57–58.**
— heat, 141–7.
— skeletons, 183–90.
— starch, 78.
annual plants, 216.
†anode, 352.
†anthracene, 62.
†anthracite, 59.
anti-friction alloys, 42.
†antimony, metal, 9.
apparent time, 319.
†approach grafting, 236.
†aqueous humour, 200.
arc, electric, 351.
— welding, 351–2.
†armature, **374**, 380–1.
artificial selection, 259.
†assimilation, 109, 118.
†asteroid, 331.
atom-splitting, 394–5.
atomic energy, 392–5.
— nucleus, 392–5.
— number, 393.
— theory, 1–2.
— weights, 8, **10–11**, 398.
atoms, **1–2**, 392.
Avogadro's Law, 14–15.

bacteria, 108.
†baking powder, 86.
†— soda, 38–39.
balance wheel, 322–3.
ball bearings, 272.
ball-and-socket joints 184–5.
bark (bot.), **105**, 220.
basic energy requirement, 149–50.
bast, 103, 105, **219–24.**

batteries, primary, 340–4.
— secondary, 344–5.
— voltaic, 340–2.
battery acid, 345.
†bauxite, 356.
bearing metal, 42.
bearings (mech.), 272.
†beer, 84.
bell transformer, 387.
†benzene, 62, **67.**
Bessemer process, 47.
bile, 113, 119, **122–3.**
— salts, 113, 115, 121, **123.**
bi-sexual organisms, 240.
†bituminous coal, 59.
†biuret test, **94–95,** 121, 124.
blast furnace, 44–45.
†blasting gelatine, 29.
bleaching, 35.
— powder, 40.
blood, 163–5.
— arterial, 165.
— clotting of, 163–5.
— de-fibrinated, 164.
— de-oxygenated, 165.
— fibrin, 164.
— film, 164–5.
— oxalated, 164.
— oxygenated, 165.
— plasma, 165.
— serum, 164.
— venous, 165.
Board of Trade unit, 368.
body (astron.), 324–32.
— (mech.), 260.
— temperature, 141–7.
bone black, 58.
— charcoal, 55, **57–58.**
— oil, 57.
boro-silicate glass, 49–50.
†boron, 49.
bottle glass, 49.
bottled gas, 66.
brain, 196–7.
brake, 272, 289, **296.**
— horse-power, 289–90.
†brandy, 84.
brass, 42.
brick, 49–50.
†bronze, 42.
brown coal, 59.
brush (elect.), 374–5.
budding, of Hydra, 239.
bud-grafting, 235–6.
Bunsen flame, 71–72.
butane, 66.
†butyric acid, 87.

Calcium, 41.
†— aluminate, 51.
— carbide, 25, 67.
†— cyanamide, 25.
— nitrate, 26.
†— oxalate, 164.
— phosphate, 52.
†— silicate, 51.
calorific value of foods, 152.
†cambium, 219–24.
candle flame, 70–71.
cane sugar, 74, **75–77.**
†canine (teeth), 189–90.
†capsule, of joint, 184.
carbohydrates, **74–81,** 152–3.
carbohydrate-splitting enzymes, 111.
†carbolic acid, 62.
carbon, 17, **55–59.**
— dioxide, 4, 6, 10.
— filament bulb, 349.
— microphone, 389.
†— monoxide, 4, 6, 10, 17, 44, 61, **62–63.**
— silicide, 52.
†carborundum, 52.
†cartilage, 184.
†casein, 121.
cast iron, 45–46.
†Castner-Kellner process, 39–40.
catalysts, 26, 37, **110–11.**
†cathode, 352.
caustic potash, 38–40.
— soda, 39–40.
†celestial pole, 324–5.
cells, 97–100.
cell division, 252–6.
— nucleus, 98–99.
— sap, 98–99.
— wall, **97–99,** 107.
cellulose, 74, **78–79,** 98, 107.
cement, 49, **51–52.**
— hydraulic, 51.
— Portland, 51.
— of tooth, 188.
central canal, 196.
— cylinder, 223.
— nervous system, 196–7.
centre of gravity, 282–4.
— of mass, 282–4.
cerebellum, 197.
cerebral hemispheres, 196–7.
cerebrum, 196–7.
†cerium oxide, 72.
C.G.S. system, 267.
chain reaction, 394–5.
changes, chemical, 12–14.
charcoal, animal, 55, **57–58.**
— bone, 55, **57–58.**
— decolorizing, 57–58.
— wood, 55–57.
chemical changes, 12–14.
— combination, 1–8.
— composition, 2–3.
— compounds, 2–17.
— control, 210–12.
— effects of electric currents, 352–8.
— energy, 294–5, 343.
— equations, 12–14.
— formulae, 9–10.
— laws, 2–16.
— messengers, 210–12.
— reactions, 12–14.
— symbols, 9.
†chitin, 168, 183.

chlorophyll, 102–3.
choke (elect.), 386–7.
†choroid, 201.
†chromosome, 252–6.
— number, 253.
†citric acid, 86.
civil year, 317.
clay, 49.
†clinometer, 286.
clotting of blood, 163–5.
— of milk, 121–2.
coal, 59.
— bituminous, 59.
— fossils, 59.
— gas, 59–62.
— hard, 59.
— soft, 59.
— steam, 59.
— tar, 61–62.
cochlea, 203–4.
cockroach, excretion in, 168.
— movement in, 183.
— nervous system of, 195.
— nutrition in, 129–30.
— reproduction in, 244–5.
— respiration in, 135.
†cod-liver oil, 87.
cohesion, 394.
coiled coil filament, 350.
†coke, 44, 55, **60–61,** 63.
cold-blooded animals, 141–2.
†colloid, 78, **118,** 125–6.
colon, 116.
coloured glass, 50.
combination, chemical, 1–8.
combining volumes, 14–15.
— weights, 5–8.
combustion and flame, 69–74.
†comet, 327, **331.**
†commutator, **374–5,** 380–1.
complex sugars, 74–77.
complex carbohydrates, 74, **77–81.**
component, of force, 275.
— of velocity, 261.
composition of forces, 275–9.
— of velocities, 261–2.
— of water by weight, 18–20.
compound pendulum, 323.
— pulleys, 304–8.
— screws, 316.
— wheel and axle, 310–11.
compounds, chemical, 2–17.
compression waves, 274.
concrete, 51–52.
— reinforced, 51.
conductors, electrical, 335, 347, **396–7.**
conservation of energy, 2, **287–92,** 295–7.
— of matter, **2–3,** 147, 394.
constant composition, law of, 2–3.
— temperature animals, 141–2.
†constellation, 324.
contact process, 37–38.
continuous currents, 339, **378.**
control experiments, **79–80,** 95, 160, 178.
†co-ordination, 194.
copper, 41, 42, **356.**
— oxide, 5, **18.**
— plating, 354.
†cordite, 29.
†cork-oak, 97.
†cornea, 198–201.
†corona (astron.), 328.
corrosion of metals, 41–42.
†coulomb, 357.
†cracking heavy oils, 69.
crop, of bird, 127.
— of cockroach, 129.
cross-fertilization, 240.
†crowbar, 299.
crown (of tooth), 188.
crucible steel, 46.
†cryolite, 356–7.
†crystalloid, **118,** 125–6.
†cupric oxide, **5,** 18.
†cuprous oxide, 5.
†curds, of milk, 122.
current, electricity, 339–46.
— interrupter, 381.
— reverser, **374–5,** 380–1.
cut glass, 49.
†cytoplasm, 98.

Dalton's atomic theory **1–2,** 395.
Davy lamp, 72–74.
D.C., 380–1.
dental formula, 190.
dentine, 187–8.
de-polarization, 341–2.
†destructive distillation, **55,** 60.
development of chick, 248–9.
†dextrin, 78, **79–80,** 112, 120.
dextrose, 74–75.
†diagonal, 261.
†dialysis, 117–18, **125–6.**
diamond, 55, **58–59.**
diaphragm, 140.
†diastase, **76,** 105, 107.
diet, 151–3.
†differential, 311.
differential pulley block, 312–13.
— respirometer, 159–61.
— screw jack, 316.
— wheel and axle, 310–11.
†differentiation of cells, 215.
diffusion, **116–18,** 125–6, 153–5.
digestion, 109, **110–25.**
— in Amoeba, 131–2.
— in bird, 126–7.

digestion in cockroach, 129–30.
— in earthworm, 130.
— in fish, 128.
— in frog, 127.
— in Hydra, 130–1.
— in mammal, 110–25.
— in Paramecium, 131.
— in snail, 128–9.
direct current, 380–1.
†di-saccharides, 74–77.
discharge lamps, 350.
†dislocation, 184.
†dissociation, electrolytic, 352–4.
distillation, destructive, **55,** 60.
— fractional, 61–62, **68–69,** 82.
dominant genes, 258.
dry cell, 343.
dry weight, increase in, 216, **225–6.**
†ductility, 46.
†ductless gland, 211–12.
†duodenum, 113.
†Duralumin alloy, 42.
†dynamite, 29.
dynamo, 376, **378–81,** 397.

ear, 201–6.
— case, 201.
— drum, 201–2.
— external, 201.
— inner, 202–6.
— internal, 201–2.
— middle, 202–6.
Earth, in solar system, 324–31.
earthworm, digestion in, 130.
— excretion in, 168.
— movement in, 181.
— nervous system of, 195.
— reproduction in, 240–3.
— respiration in, 134.
†ebonite, 32.
efficiency (mech.), 307.
egg-cell, 236.
Einstein, 394.
elastic bodies, 270.
— limit, 274.
elasticity, 273–4.
electric arc, 351.
— charges, **333–9,** 396–7.
— currents, **339–46,** 397.
— energy, 384–7.
— generators, 378.
— lighting, 349–52.
— motors, 374–6.
— potential, 369.
electrical resistance, 347, **368–9,** 385, 397.
— units, 366–9.
electro-magnet, 374–8.
electro-magnetic induction, 376–8.
electro-metallurgy, 355–6.
†electro-motive force, 367.
electro-plating, 42, **354–5.**
electro-static induction, 336–8.
electro-statics, 333–9.
electrotype, 355.
electrode, 352.
electrolysis, 39–40, 44, **352–7.**
— of brine, 39–40.
†electrolyte, 352.
electron, 392–7.
†electroscope, 335–8.
†Elektron alloy, 42–43.
E.M.F., 367.
†emulsion, 91–92, **122–3,** 125.
†enamel, dental, 187–8.
†endosperm, 106.
energy, definition of, 287.
— electrical, 384–7.
— kinetic, 290–4.
— measurement of, 287.
— mechanical, 291–2.
— potential, 290–1.
energy requirement, human, 149–50.
— transformation of, 294–7.
— values of foods, 151–3.
— and matter, 287, **394.**
enzymes, 75, 76, 82, 78, 93, 107, **110–11,** 133, 141.
†eosin, 57.
epiglottis, 167.
equations, chemical, 12–14.
— of motion, 263–6.
†equilibrant, **276,** 278.
†equilibrium, 270, 282, **284–6.**
— neutral, 284–5.
— stable, 284–5.
— unstable, 284–5.
equivalence of mass and energy, 394.
equivalent, chemical, 6–8.
— of copper, 18–20.
— of oxygen, 18–20.
— weight, 2, **5–8.**
†erepsin, 114.
†escapement wheel, 322–3.
†ester, 86–87.
ethane, 66.
†ethyl acetate, 87.
†— alcohol, 81–87.
†ethylene, 66.
†Eustachian tube, 202–3.
evolution, 237, **258–9.**
excretion, in Amoeba, 168.
— in animals, 168–72.
— in bird, 171.
— in cockroach, 168.
— in earthworm, 168.
— in fish, 169.
— in frog, 169.
— in Hydra, 168.
— in mammal, 169–72.
— in Paramecium, 168.
— in plants, 172.
— in snail, 168–9.
exercise and carbon dioxide, 162–3.

exercise and pulse rate, 162–3.
expired air, 155–7.
explosives, 29.
extension of elastic bodies, 273.
external ear, 201.
— respiration, 140.
extraction of metals, 43–44.
eye, 198–201.
— ball, 198.
— dissection of, 200–1.
— lens, 201.
— lid, 198.
— muscles, 201.
— socket, 198.

faeces, 116.
falling bodies, 266–7.
Faraday, 376–8.
fatigue, muscular, 193.
fat-splitting enzymes, 111.
fats and oils, **87–92,** 152–3.
— — tests for, 88–89.
fatty acids, **86–87,** 89–91, 114–16.
Fehling's test, 5, **75,** 76, 77, 80.
†fermentation, acetic, 86.
†— alcoholic, 81–86.
†ferric sulphate, 17.
†ferrous sulphate, 17.
fertilization, 229, **236–8.**
— cross, 240.
— external, 245–6.
— internal, 244, **247–51.**
— self, 240.
†fibrin, 164.
†fibrinogen, 164.
filterable virus, 96.
†fire-damp, 73–74.
fire-fly, 351.
fish, digestion in, 128.
— excretion in, 169.
— movement in, 191.
— nervous system of, 196.
— reproduction in, 245.
fish, respiration in, 136.
†fission, nuclear, 394–5.
fixation of nitrogen, 24–26.
fixed proportions, law of, 2–3.
flame and combustion, 69–74.
†flint, 49.
— glass, 49.
†flotation—separation of ores, 43.
flowers of sulphur, 31.
†fluorescent lamp, 351.
flying, of birds, 191–2.
focal length of eye, 199–200.
focusing, in eye, 199–200.
food and energy, 147–53.
foot-pound, 287–8.
force, 269–73.
— diagrams, 274–9.
— definition of, 270.
— measurement of, 272–3.
— of gravity, 269, **271.**
— ratio, 304.
forces in equilibrium, 277–80.
fore-brain, 196–7.
forging, 46.
formulae, chemical, 9–10.
fossils, coal, 59.
†fovea, 199–200.
F.P.S. system, 317.
fractional distillation, 61–62, **68–69,** 82.
†Frasch process, 31.
fresh weight, increase in, 216, **225–6.**
friction, 268, 270, **271–2,** 296, 306.
— kinetic, 272.
— rolling, 272.
— sliding, 272.
— static, 272.
frictional electricity, 333–9.
frog, digestion in, 127.
— excretion in, 169.
— reproduction in, 245–6.
frog, respiration in, 138–9
†fructose, 74–75.
fungi, 108.
†fungicide, 32.
†fuse (elect.), 348.

'*g*', 267.
†galactose, 76.
†galaxy, 326.
Galileo, 265, 266–7, 268, **319–20.**
Galvani, 339–40.
†galvanometers, **363–6,** 378.
†gamete, 236.
gas-filled bulbs, 349–50.
†gas-mantle, 72.
†gas-oil, 69.
gastric juice, 112–13.
Gay-Lussac's law of combining volumes, 14.
†gene, 254–9.
†German silver, 42.
germination, 216.
†gestation period, 251–2.
gill, 136–7.
— cover, 136.
— raker, 128.
— slit, 136.
gizzard, of bird, 127.
— of cockroach, 129–30.
— of earthworm, 130.
glass, 49–50.
— bottle, 49.
— boro-silicate, 49–50.
— coloured, 50.
— cut, 49.
— etching, 50.
— flint, 49.
— hard, 49.
— lead, 49.
— potash, 49.
— soda, 49.
— soft, 49.
— window, 49.
glazing pottery, 50–51.
gliding, of birds, 192.

glottis, 167.
glucose, **74–75,** 114, 141.
†glyceride, 87.
†glycerine/glycerol, 86, **87–90,** 114–15.
†glycine, 93.
glycogen, 78, 80, 81, 112, **119.**
gold-leaf electroscope, 335–6.
good conductors (elect.), 335, **396–7.**
†gootee layering, 234–6.
†grafting, 233–6.
— by approach, 235–6.
Graham, 118.
†gram-equivalent, 7.
gram-molecular volume, 15–16.
gramophone records, 355.
grape sugar, 74–75.
†graphite, 55, **58.**
†gravimetric analysis, 7.
gravitation, 271.
gravity, 269, **271.**
— centre of, 282–4.
— and plant growth, 173–5.
grip rule (elect.), 362.
growth, 215–26.
— in dry weight, 216–17.
— in length, 217–18.
— in thickness, 218–26.
— measurements, 224–6.
— movements in plants, 173–80.
— rings, 221–2.
— secondary, 215–22.
†guano, 171.
guard-cells, 103–4.
†gum (of tooth), 188.
†gun-cotton, 29.
†— metal, 42.
— powder, 29.

†Haber process, 26–27.
†haemocyanin, 136.
haemoglobin, 93, 134, 141, **165.**
hair spring (of watch), 323.
†Haldane apparatus, 159.
†halo (astron.), 331.
hard coal, 59.
— glass, 49.
— soap, 89–92.
hardness of water, 90.
hearing, 201–5.
— experiments, 205.
heart, examination of, 165–7.
— muscle, 192–3.
— sac, 165.
— valves, 166–7.
heartwood, 98, **108.**
heating effect of electric current, 347–52.
heat loss by conduction, 145–6.
— — by convection, 145–6.
— — by evaporation, 143–4.
— — by radiation, 144–5.
— of combustion, 148.
— regulation, 142–7.
— treatment of steel, 48.
heavy hydrogen, 396.
†hepatic portal vein, **116,** 118.
hermaphrodite animals, 240.
hexane, 66.
†hibernation, 142.
high explosives, 29.
hind-brain, 196–7.
hinge-joints, 185.
honey, 76.
Hooke, 97.
Hooke's law, 273–4.
†hormone, 177–9, **210–12.**
†horse-power, 288–9.
hot-wire ammeter, 366.
hour-candle, 318.
hour-glass, 318.
†hybrid, 257.
Hydra, digestion in, 130–1.
— excretion in, 168.
Hydra, movement of, 180.
— nervous system of, 195.
— reproduction in, 239–40.
— respiration in, 134.
hydraulic cement, 51.
hydrocarbons, 65–68.
hydrochloric acid, 27, **40.**
†hydrofluoric acid, 50.
hydrogen, reduction with, 17.
— chloride, 6, **15,** 40.
†— fluoride, 50.
— ion concentration, 353–4.
— peroxide, 17.
— sulphide, 17, **32–34.**
†hydrolysis, **75–76,** 114.
†hydroxyl ion, 353.

ignition temperature, 53, **72–73.**
incisor, 189–90.
inclined forces, 275–8.
— plane, 313–14.
indestructibility of matter, **2–3,** 394.
induced charges, 336–8.
— currents, 378.
induction coil, 381.
— electro-magnetic, 376–8.
— electro-static, 336–8.
†inertia of matter, 269.
†ingestion, 109–10.
inheritance, 256–9.
inner ear, 202–6.
inspired air, 155–7.
instinct, 212–14.
insulators (elect.), 335, **396.**
intelligence, 213–14.
internal ear, 201–2.
— fertilization, 244, 247–51.
— respiration, **133,** 141.
— sense-organs, 206.
intestinal juice, 114.
Invar steel, 43.

†invert sugar, 76.
involuntary muscle, 192–3.
†iodoform, **83,** 86.
†ion, **352–4,** 393.
†ionization, 352–4.
iris, of eye, 199–201.
iron, 41–42.
— and steel, 44–49.
†— disulphide, 31.
†— pyrites, 31.
— sulphide, 31, **32–33.**
irritability, 194.
isotopes, 395–6.
†ivory, 188.

jam, 76.
jet-propulsion, 271.
joints, skeletal, 184–5.
Jupiter, **328–9,** 331.

†katabolic changes, 101.
kerosene, 69.
kidney, 119, **168–72.**
— examination of, 171–2.
kilo-calorie, 149.
†kilowatt, 289, **368.**
kilowatt-hour, 368.
†kinetic energy, **290–4,** 296.
†— friction, 272.
Kipp's apparatus, 33.

†lactase, 114.
†lacteal, 115–16.
†lactose, 76–77.
†lamp black, 55, **57.**
large intestine, 116.
†latitude, 319.
law of combining weights, 2–6.
— of conservation of energy, 2, **147,** 295–7, 343, 394.
— of conservation of matter, **2–3,** 394.
— constant composition, 2–3.
— of equivalents, 2–6.
law of fixed proportions, 2–3.
— of indestructibility of matter, **2–3,** 394.
— multiple proportions, 2–5.
— reciprocal proportions, 2–6.
laws of chemical combination, 2–16.
— of motion, 268–70.
Lavoisier, 2.
†layering, 234–6.
†leaching, of ores, 43.
lead accumulator, 344–5.
— chamber process, 36–37.
— glass, 49.
— nitrate, 28.
— oxide, 16.
†— peroxide, **17,** 53.
leaf mosaic, 179.
leap years, 317.
Leblanc process, 38.
†Leclanché cell, 342.
leverage, 281–2.
levers, 299–303.
— Class I, 299–300.
— Class II, 301.
— Class III, 301–2.
— in animal body, 302.
lifting tackle, 303–8.
light alloys, 42–43.
— and plant growth, 176–9.
— energy, 294–5.
†— year, 326.
lightning, 338–9.
— arrestor / conductor, 338–9.
†lignin, 107.
†lignite, 59.
limb girdles, 183–4.
†lipase, 107, 111, 113, **124–5.**
liver, 116, **118–19.**
local action in voltaic cells, 340.
— apparent time, 319.
long-distance transmission (elect.), 385–6.
longitudinal waves, 274.
lungs, examination of, 165–7.
— of bird, 139–40.
— of frog, 138.
lymph, 134, 138.
— hearts, 138–9.
— vessels, **115–16,** 138.

machines, 298–316.
— efficiency of, 296, 307.
†macula lutea, 200.
†Magnalium alloy, 42.
magnesium, 41.
— manufacture of, 357.
magnetic effect of electric current, 358–62.
— field round conductors, 360–2.
†magneto, 381.
magneto-electric effect, 378–9.
make-and-break, 381–2.
†malleability, 46.
†malt, 76.
— sugar, **76–77,** 112.
†maltase, 76, **114.**
†maltose, **76–77,** 112.
mammals, digestion in, 108, 110–25.
— excretion in, 169–72.
— movement in, 183–6.
— nervous system of, 196–8.
— reproduction in, 249–52.
— respiration in, 140–1.
manganese dioxide, 17.
†mantle, gas, 72.
†marcottage, 236.
margin of safety (mech.), 286.
Mars, 328–31.
†marsh gas, 65–66.
mass (mech.), 271, **272,** 394.

mass and energy, equivalence of, 394.
— centre of, 282–4.
matches, 53–54.
matter and energy, 287, **394.**
measurement of electricity, 363–72.
— of energy, 287.
— of force, 272–3.
— of growth, 224–6.
— of power, **288–9,** 367–8.
— of time, 317–23.
— of work, 287–8.
†mechanical advantage, 304–16.
— efficiency, 296, **307.**
— energy, 291–2.
†medulla oblongata, 197.
†— of kidney, 170.
Mendel, 256–9.
Mercury, 328–9.
†metabolism, 101.
†metallurgy, 43–48.
metals, characteristics of, 40–43.
— extraction of, 43–49.
metamorphosis, 183.
meteors, 327, **331–2.**
methane, 6, 55, **65–66,** 68.
†methyl alcohol, **55,** 57, 82.
methylated spirit, 82.
†methylene blue, 57.
†microphone carbon, 389.
mid-brain, 196–7.
middle ear, 202–6.
mild steel, 48.
milk sugar, 76–77.
milli-ammeter, 369–70.
milli-ampere, 369.
milli-volt, 370.
milli-voltmeter, 370.
mineral oil, 68–69.
— ores, 43.
miner's safety lamp, 72–74.
mining, 43.
†molar (tooth), 189–90.
†molasses, 75.
molecular weight, **12,** 15–16.
molecules, 1, **8–9,** 392.
molluscs, *see* snail.
†moment (mech.), 281.
monoclinic (β) sulphur, 32.
†mono-saccharides, 74–75.
Moon, 325–6, **330–1.**
Morse Code, 388.
motion, 260.
— equations of, 263–6.
— laws of, 268–70.
motor, electric, 374–6.
movement in animals, 180–94.
movements of plants, 173–80.
moving-coil galvanometer, 365–6, **369–72.**
— iron galvanometer, 363–4.
— magnet galvanometer, 363–4.
multiple proportions, law of, 2–5.
muscle, 185–6, **192–4.**
— cells, 192.
— fibres, 192.
— heart/cardiac, 192–3.
— involuntary / smooth / unstriped, 192–3.
— voluntary/striped/skeletal, 192–3.
†mutton-fat, 87.

naphthalene, 62.
native metals, 43.
†natural gas, 65–66.
— selection, 259.
neck (of tooth), 188.
needle (β) sulphur, 32.
negative electricity, 334–8.
Neptune, 328–9.
nervous system, in Amoeba, 194.
nervous system in earthworm, 195.
— — in Hydra, 195.
— — in insects, 195.
— — in mammals, 196–8.
— — in molluscs, 196.
— — in Paramecium, 194.
neutral equilibrium, 284–6.
neutrons, 393–6.
Newton, 268–71.
Newton's laws of motion 268–71.
Nichrome alloy, **43,** 348.
nitric acid, 17, **26–28.**
†— anhydride, 28.
†— oxide, 25–26, **28,** 37.
nitro-compounds, 29.
†nitro-glycerine, 29.
nitrogen compounds, 21–29.
— cycle, 24.
†— dioxide, 28.
nitrogen-fixing bacteria, 24–25.
†nitrogen monoxide, 28.
— oxides, 23, **28–29.**
†— pentoxide, 28.
†— peroxide, 25–26, **28,** 37.
†— tetroxide, 28.
†— trioxide, 28.
†nitrous anhydride, 28.
†— oxide, 28.
†nodule bacteria, 25.
non-conductors (elect.), 335.
non-electrolytes, 352.
non-metals, characteristics of, 40–41.
normal sodium sulphate, 27.
— solutions, 7–8.
Norwegian saltpetre, 26.
nuclear chain reaction, 394–5.
— energy, 393–6.
†— fission, 394–5.
nucleus, of cell, 98–99.

nutrition, in Amoeba, 131–2.
— in bird, 126–7.
— in cockroach, 129–30.
— in earthworm, 130.
— in fish, 128.
— in frog, 127.
— in green plants, 102–8.
— in Hydra, 130–1.
— in mammals, 108–26.
— in non-green plants, 108.
— in Paramecium, 131–2.
— in snail, 128–9.

Oerstedt, 359.
†ohm, 368–9.
Ohm's law, 368.
oils and fats, **87–92,** 152–3.
— — tests for, 88–89.
†oleic acid, 87.
†olive-oil, 87.
open-hearth process, 47–48.
optic nerve, 201.
†orbit (astron.), 317.
†ore-dressing, 43.
†ore-flotation, 43.
ore-leaching, 43–44.
ore-minerals, 43.
ores, 43.
organ of balance, 203–6.
— of hearing, 202–4.
— of position, 203–6.
— of smell, 208–9.
— of taste, 206–8.
— of touch, 209–10.
organic chemistry, 55.
osmosis, 117–18.
oval window (of ear), 203.
ovary, 238, **251.**
overshot water-wheel, 293.
†ovo-testis, 243.
ovum, 236.
oxidation, 16–17.
oxides of nitrogen, 23, **28–29.**
oxidizing agents, 16–17.
†oxy-acetylene flame, 67.
oxy-haemocyanin, 136.
oxy-haemoglobin, 134, 141, **165.**

†palisade cells / tissue, 103–4.
†palmitic acid, 87.
pancreas, 113.
pancreatic juice, 113, **123–5.**
panting, 144.
paraffin series, **66,** 68.
— wax, 69.
parallel connection (elect.), 371, **372–3.**
— forces, 280–1.
parallelogram of forces, 276–7.
— of velocities, 261.
Paramecium, excretion in, 168.
— movement in, 180.
— nutrition in, 131–2.
— reproduction in, 238–9.
— respiration in, 134.
— sensitivity in, 194.
parasites, 96, **108.**
parental care, **249–50,** 252.
†peat, 59.
†Pelton wheel, 293–4.
pencils, lead, 58.
pendulum, 292, **319–23.**
penis, 243, **251.**
pentane, 66.
†pepsin, 112, 120–1.
†peptones, **94–95,** 124.
periodic table of elements, 393.
perpetual motion, 269.
petrol, 69.
petroleum, 68–69.
— jelly, 69.
— refining, 68–69.
†pewter, 42.
pH scale, 353–4.
phenol, 62.
†phenolphthalein, 124.
†phloem, 103, 105, **219–24.**
phosphate rock, 52, **171.**
phosphates, 54.
phosphoric acid, 52, **54.**
†phosphors, 351.
phosphorus, allotropic forms, 52–53.
— cycle in Nature, 171.
— manufacture of, 52.
— pentoxide, 52–54.
— red, 52–54.
— white/yellow, 52–54.
physiology, 101.
pigeon, *see* bird.
pig iron, 45–46.
†pitch (from tar), **62,** 69.
†— of screw, 315.
pith-ball electroscope, 335.
†pituitary gland, 211.
pivot, 299.
plain muscle, 192–3.
†planetoids, 327, **331.**
planets, 325–6, **328–30,** 332.
plant hormones, 177–9.
plasma, 165.
plastic sulphur, 32.
plastids, 77–78, 100, **102–3.**
Pluto, 328–9.
†polarization (elect.), 341–2.
polygon of forces, 278.
†poly-peptides, 94, 124.
poly-saccharides, 77–81.
†portal vein, **116,** 118.
Portland cement, 51.
positive electricity, 334–8.
†posture, 197.
potash glass, 49.
potassium, 41.
†— dichromate, 17.
— hydroxide, 38–40.
— nitrate, 17.
†— oxalate, 164.
†— permanganate, 17.
†potential difference, 367.

†potential energy, **290–1**, 296.
pottery, 49–50.
power (mech.), 288–9.
— units of, 288–9.
†pre-molar (tooth), 189.
primary cell/battery, 340–2.
— coil, 377, 381, **383–4**.
principle of work, **306**, 314, 315.
†producer-gas, 48, 61, **63–65**.
propane, 66.
propellent explosives, 29.
†proteases, 107, **111**.
proteins, 24, 32, 33, 52, **92–96**, 124, 152–3.
— tests for, 94–95.
protein-splitting enzymes, 111.
†proteoses, **94–95**, 121, 124.
protons, 393–6
protoplasm, 98–99.
Proust, 3.
†ptyalin, 76, 81, 112, **119–20**.
†pulleys, 303–13.
pulp cavity (of tooth), 187–8.
pulse-rate and exercise, 162–3.
pupil (of eye), 199–201.
†pylorus, 113.
†pyrogallol, 156–7.

qualitative analysis, 34.
quantitative analysis, 7.
quantity of electricity, 357.
quartz, 49.
†quenching, of metals, 48.

reacting quantities, **12–14**, 16.
reactions, chemical, 12–14.
reasoning, 213–14.
receiver, telephone, 388–90.
†receptor, 198–210.
†recessive gene, 258.
†rectifier, copper oxide, 387.
†— rotary, 387.
†— valve, 387.
rectum, 116.
reducing agents, 16–17.
†reduction (chem.), 16–17.
— division, of cells, 254–6.
†reinforced concrete, 51.
†rennin, 112.
†resolution of forces, 275, **279–80**.
resultant force, 275–9.
— velocity, 261.
retina, 199–201.
reversible reactions, 13.
rhombic (α) sulphur, 31–32.
ring-barking, 105.
†Rochelle salt, 75
rock oil, 68–69.
— phosphate, 52, **171**.
rocket propulsion, 271.
roll sulphur, 31.
roller bearings, 272.
root (of tooth), 188.
rotary converter, 387.
round window (of ear), 203.
rubber, 57.
— vulcanized, 32.
†rum, 84.
rusting, of iron, 41–42.
— of other metals, 41–42.

safety lamp, 72–74.
— matches, 53–54.
saliva, 81, 109, **111–12**, 119–20.
sand, 49.
— glass, 318.
sap, 98–99.
†saprophyte, 108.
sap-wood, 108.
†satellites, 325–6.
Saturn, 328–9.
†sclerotic, 201.
screw, 314–16.
— compound/differential, 316.
†screw-jack, 315–16.
screw rule (elect.), 346.
screw-thread, 314.
secondary cell / battery, 344–5.
— coil, 377, 381, 383–4.
— growth/thickening, 215–22.
see-saw levers, 299.
self-fertilization, **240**, 256–7.
†semicircular canals, 205–6.
sense of balance and position, 205–6.
— of feeling, 209–10.
— of hearing, 201–5.
— of pain, 209–10.
— of sight, 198–201.
— of smell, 206–9.
— of taste, 206–9.
— of temperature, 209–10.
— of touch, 209–10.
sense-organs, 198–210.
— internal, 206.
sensitivity, 194.
series connection (elect.), 370, 372–3.
series-wound motor, 376.
†serum, 164–5.
†Shaddock/Pomelo, 236.
short circuit, 348.
†shunt (elect.), 371.
†shunt-wound motor, 376.
silica, 49.
silicon, 49–52.
†— carbide, 52.
simple cell, 340–2.
— sugars, 74–75.
skeletons, 182–90.
— animal, 183–90.
— external, 183.

skeletons, internal, 183–90.
skin, 209.
— skeleton, 186–7.
†slag, 44–45.
slip rings (elect.), 378–9.
smell-patch, 208–9.
†smelting, 44–45.
smooth muscle, 193.
snail, excretion in, 168–9.
— movement in, 182.
— nervous system of, 196.
— nutrition in, 128–9.
— reproduction in, 243–4.
— respiration in, 135–6.
soap, 89–92.
— hard, 89–92.
— soft, 89–92.
soda ash, **39**, 49.
— glass, 49.
sodium, 41.
†— acetate, 89.
— amalgam, 40.
†— bicarbonate, 38–39.
— carbonate, **38–39**, 49.
— chloride, 6, **39–40.**
†— cyanide, 44.
†— hydride, 6.
— hydrogen sulphate, 27.
†— hydrogen tartrate, 75.
†— hydrosulphite, 165.
— hydroxide, 39–40.
— nitrate, 27.
†— silicate, 49.
— sulphate, 27.
soft coal, 59.
†solar day, 317.
— energy, 296–7.
†— system, 327–8.
— time, 319.
solder, 42.
†solenoid, 360–1.
Solvay's process, 38–39.
soot, 55, **57.**
†Sorghum, sweet, 75.
†spark plug, 382.
speech, 214.
speed and velocity, 260–1.
sperm-sac/receptacle, 242–3.
sperms, **236–8**, 240–7.
spinal bulb, 197.
— cord, 196–7.
spirits, alcoholic, 84.
— methylated, 82.
splitting atoms, 394–5.
†stability (mech.), 284–6.
stable equilibrium, 284–6.
standard gas-volume, 15–16.
— solutions, 7–8.
— time, 317.
starch, 77–81.
static electricity, 333–9.
†— friction, 272.
steam coal, 59.
†stearic acid, 87.
steel and iron, 44–49.
— crucible, 46.
— mild, 48.
— structural, 48.
— tempered, 48–49.
— tool, 48.
†steelyard, 299.
step-down transformer, 383–4.
step-up transformer, 383–4.
stimulus, 173.
stinging-animals, *see* Hydra.
†stock (grafting), 233–4.
storage cells, 344–5.
strain (mech.), 270.
striped muscle, 192–3.
structural steel, 48.
struggle for existence/survival, 258–9.
†sucrase, 114.
†sucrose, 74, **75–77**, 112.
Sudan III test for fat, 88–89.
†sugar-beet, 75.
†sugar-cane, 75.
†sugar-palm, 75.
sugars, simple and complex, 74–77.
sulphur, 29–32.
— allotropes of, 31–32.
— α (rhombic), 31–32.
— β (monoclinic), 32.
— dioxide, 4–5, 17, **34–36.**
— extraction of, 30–31.
†— trioxide, 4–5, **36–38.**
— uses of, 32.
†sulphuretted hydrogen, 17, **32.**
sulphuric acid, 5, 35, **36–38.**
†sulphurous acid, 5, **35.**
Sun, 327–8.
†sun-dial, 318.
surface tension, 113, 122–3.
sweating, **143–4**, 172.
swimming, of fish, 190–1.
— rule (elect.), 359.
symbols, chemical, 9.
†sympathetic nervous system, 197–8.
†synovial fluid, 184.
synthetic ammonia, 26–27.

†tackle, lifting, 305.
tapping-key, 388.
†tartaric acid, 86.
taste, sense of, 206–9.
†— bud, 207.
— organ of, 206–9.
tear gland, 198.
tears, 198.
teeth, 187–90.
— milk/temporary, 187–90.
— permanent, 187–90.
telegraph, 388.
telephone, 388–91.
— receiver, 388–90.
— transmitter/microphone, 388–90.
†teleprinter, 388.
†tempering steel, 48–49.
tension (mech.), 270.
tentacles, of Hydra, 130.
— or horns, of snail, 242.
†terminal (elect.), 345.

testes, 238–50.
†thorium oxide, 72.
thrust (mech.), 270.
thunder, 338.
†thyroid gland, 211.
†tissue respiration, **133**, 141.
tissues, 97.
†tobacco mosaic disease, 95.
†toluene, 67.
tool-steel, 48.
transformations of energy, 294–7.
transformer (elect.), 382–6.
transmitter, telephone, 388–90.
transverse waves, 274.
†treacle, 75.
triangle of forces, 277–8.
†tri-nitro-toluene (T.N.T.), 29.
†trypsin, 113.
†tungsten-filament bulb, 349.
turbine, steam, 295.
— water, 293–4.
†tusks, 190.
type-metal, 42.

ultra-violet light/rays, 351.
undershot water-wheels, 293.
uniform acceleration, 262–4.
— velocity, 260–1.
unstable equilibrium, 284–6.
unstriped muscle, 192–3.
uranium, isotopes of, 396.
Uranus, 328–9.
urea, **119**, 172.
†uric acid, 171.
urine, **119**, 170.
uterus, 251.

vagina, 243.
†valency, 8.
valve rectifier/converter, 387.
†vanadium pentoxide, 37.
vapour density, 16.
variable-temperature animals, 141–2.
— velocity/speed, 260–7.
vascular bundles, 103.
vaseline/petroleum jelly, 69.
†velocity and speed, 260–7.
— ratio (mech.), 306–14.
ventral nerve cord, 195.
Venus, 328–9.
vertebrate animals, 183–6.
†villus, 115.
†vinegar, 86.
virus, 96.
†— proteins, 95–96.
†visual purple, 200.
†vitreous humour, 200.
vocal cords, 167.
volcanic sulphur, 29–31.
†volt, 367, **369**.
Volta, 340.
†voltage, 350.
†voltaic cell/battery, 340–3.
†voltmeter, 369–72.
†volumetric analysis, 7–8.
— solutions, 7–8.
voluntary muscle, 192–3.
†vulcanite, 32.
†vulcanized rubber, 32.

walking movements, 191.
warm-blooded animals, 141–2.
washing-soda, 38–39.
water and direction of plant growth, 179–80.
— gravimetric composition of, 18–20.
water-clocks, 318.
†water-gas, 61, **63–65**.
water-glass, 49.
water-power, 290–1, **292–4**.
water-turbines, 293–4.
water-wheels, 292–3.
Watt, 288–9.
†watt (elect.), 289, 350, **368**, 384.
waves, compression/longitudinal, 274.
— transverse, 274.
wedge, principle of, 314.
weight and mass, 267, **271**.
welding, **46**, 67.
wheel-and-axle, 308–11.
†wheel-barrow, 300.
†whey, 121.
†whiskey, 84.
white phosphorus, 52–54.
wilting, 179.
window glass, 49.
wine, 84.
womb, 251.
wood, 103, 105, **219–24**.
†— alcohol, 55, **57**, 82.
— charcoal, 55–57.
†— gas, 55, **57**.
†— tar, 55, **57**.
work (mech.), 287–9.
— principle of, 306–7.
— units of, 287–9.
†wrought iron, 46.

xantho-proteic test, 95.
†xylem 103, 105, **219–24**.
†xylene, 67.

yeast, 81–85.
yellow phosphorus, 52–54.
†— spot (of eye), 200.
yolk (of egg), 249.

†zygote, 236.
†zymase, 82.

6605

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